FUNDAMENTALS OF PHYSICS

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PREFACE

This book covers the syllabuses in physics of the **All-India Certificate Course (N.C.C.)** in Civil, Mechanical and Electrical Engineering, as also of the **L.C.E., L.M.E.** and **L.E.E.** Examinations of the State Council for Engineering and Technical Education, West Bengal. In addition, it provides an adequate course in physics covering one year of study as in a **University Preparatory Course.**

The authors' aim has been to give a clear and systematic presentation of the fundamental principles of physics in all its branches. These have been amply illustrated with examples from everyday life and industry. Laboratory work for students has been put in a form appropriate to a laboratory manual. This will enable the students to work successfully in the laboratory with very little additional guidance.

Our thanks are due to Sri N. N. Sen Gupta, Principal, Calcutta Technical School, for his valuable assistance with the book as a whole, and to Sri Amiya Ranjan Mukherjee of Messrs A. Mukherjee & Co., Private Ltd., for permission to use material from the senior author's book *An Introduction to Physics.*

AUTHORS

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CHAPTER I

UNITS AND MEASUREMENTS

1. Science and its Branches

Man has always sought to know and understand all that he sees around him. His efforts, extended over centuries, have built up a vast store of knowledge in various fields. The branch of human knowledge which deals with the changes and properties of living and non-living things present in Nature is called *Natural Science.* It has two broad divisions, *biological* and *physical.* Biological sciences deal with living things; and physical sciences, with non-living things. *Physics* and *chemistry* are the two main branches of the latter. Biology, physics, chemistry and the abstract science of mathematics are the fundamental among sciences. On them are built the applied sciences, such as engineering, agriculture, aeronautics and a host of others.

2. Matter and Energy

The objects that we see all around us, e.g., books, tables, trees, stars etc. are called *bodies. Matter* is the material or stuff of which these bodies are made. Anything which has weight and occupies space is supposed to be made of matter.

The changes that we see in Nature may be classified as a change in (i) the position, (ii) the condition or (iii) the type of matter. A moving train shows a change of position. A boiling kettle exhibits a change of condition of water, from liquid to vapour. In burning coal we get an example of a change in which matter changes type—from coal to the gas carbon dioxide.

Matter cannot of itself undergo any change ; change of any kind in matter must be brought about by some agency. The invisible agency which does it is called *energy.* All natural phenomena are effects of the action of energy on matter. Matter and energy are all that exist in the physical world.

It will be of help to accept the following definitions :

Matter is anything which offers resistance when attempt is made to change its condition of rest or motion.

Energy is that which causes a change in the position, the condition or the type of matter.

A body is any piece of matter.

3. What is Physics

Matter has been found to consist of about 90 elementary types called *elements.* Water is made up of two elementary types of matter, viz., hydrogen and oxygen. Air is a mixture, chiefly of two elements—nitrogen and oxygen.

We may say that *physics deals with those phenomena of the inanimate world in which the type of matter does not change,* while chemistry is essentially concerned with phenomena in which the type changes. This distinction cannot, however, always be maintained.

Branches of Physics. Physics has six main divisions, viz., *mechanics*, *heat*, *sound*, *light*, *magnetism* and *electricity*. Of all these mechanics is the most fundamental. It deals with properties of matter, and treats of such ideas as motion, force, energy and their relations to one other and to matter. Heat, sound and light classify natural phenomena perceived through our senses of touch, hearing and vision. We have no special senses for detecting electricity and magnetism. They are studied through the effects they produce.

Physics is built on accurate measurement. In the language of physics the term *'physical quantity'*, or sometimes simply 'quantity', is used to mean anything that is capable of being measured. Length, weight, volume, time, speed, temperature etc. are common examples.

Physics seeks to acquire knowledge of natural phenomena through accurate measurement of the quantities involved. It is, therefore, called an *exact science* ; and so is chemistry. Measurement is the foundation on which physics is built.

To measure a quantity we compare it with a quantity of the same kind, which we call the *unit*. To find the length of a table we may use a foot-rule and see how many times the length known as the *foot* is contained in the length of the table. If by laying the foot-rule alongside the table we find that the latter is five times as long as the former, we say that the length of the table is 'five feet'. A measurement involves two things, a 'number' and a 'unit'. Mere mention of the number carries no sense ; the unit must be stated.

4. Fundamental Quantities and their Units

Among the quantities which physics deals with, three are regarded as fundamental. They are *length*, *mass* and *time*. Length and time are well known concepts and cannot be explained in simpler terms. *Mass is the amount of matter contained in a body.* The units of these three quantities are called the *fundamental units.*

All quantities with which we are concerned in mechanics can be defined in terms of length, mass and time. Their units can also be derived from the units of length, mass and time. For this reason, these quantities and their units are called 'fundamental'. Units of the other quantities are called 'derived' units.

5. Systems of Units

There are two systems of units in general use—(i) the C.G.S. or the Metric system, and (ii) the F.P.S. or the British system.

(i) In the C.G.S. system the fundamental units of length, mass and time are the *centimetre*, the *gram* and the *second*. This system is used for scientific measurements all the world over. It has the advantage that units of the same kind differ by multiples of ten. Besides, the units of length and mass are related to naturally occurring quantities and to themselves (see below).

(ii) In the F.P.S. system, the *foot*, the *pound* and the *second* are respectively the units of length, mass and time. It is generally used in the English speaking countries for commercial and engineering purposes. India has recently adopted the C.G.S. system for all purposes.

6. Units of Length

(*i*) *C.G.S. system*—The **Centimetre** is one hundredth part of a metre. The metre was intended to be a ten-millionth part of the length of a quadrant of the earth's circumference measured from the equator to the North Pole through Paris. But this intention was not quite fulfilled. The bar constructed on the above basis fell slightly short of the hypothetical value. For legal and scientific purposes the **metre** is now defined as the distance at 0°C between two marks engraved upon the central flat portion of a platinum-iridium bar, called the International Prototype Metre (Fig. 1), kept in an underground vault at Sevres near Paris. Several countries possess very accurate copies of this bar. India will soon have one.

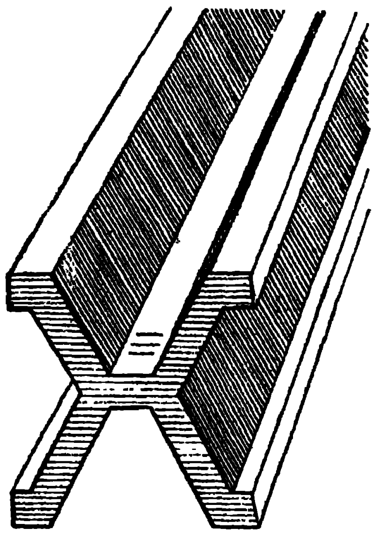

Fig. 1

(*ii*) *F.P.S. system*—The **foot** is one third the distance at 62°F between the central lines on two gold plugs in a bronze bar called the Imperial Standard Yard. It is preserved in the Standards Office, Westminister, London. In the U.S.A. the legal yard has been defined by the Congress as $\frac{3600}{3937}$ metre.

Multiples and Sub-multiples of Length

C.G.S. Units	F.P.S. Units
1 kilometre (km) = 1000 metres	1 mile (mi) = 1760 yards
1 metre (m) = 100 centimetres	1 yard (yd) = 3 feet
1 centimetre (cm) = 10 millimetres (mm)	1 foot (ft) = 12 inches (in)

Conversion Factors for Length

To convert lengths from one system to the other, remember

1 metre = 39·37 inches
1 inch = 2·54 centimetres

7. Units of Mass

(*i*) *C.G.S. system*— The **gram** is one-thousandth part of the mass of a certain platinum-iridium cylinder called the International Prototype Kilogram and kept at the same place as the standard metre. It was intended that the weight of the cylinder should be equal to that of one thousand cubic centimetres of pure water at its maximum density (i.e., at 4°C). The intention was not quite fulfilled. The agreement, however, is close enough. For all practical purposes we may take it that *one gram is the mass of one cubic centimetre of water.*

(*ii*) *F.P.S. system*—The **pound** is defined as the mass of a certain platinum cylinder marked 'P.S. 1844, 1 lb' and preserved at the same place as the Standard Yard.

Multiples and Sub-multiples of the Units of Mass

C.G.S. Units	F.P.S. Units
1 kilogram (kg) = 1000 grams	1 ton = 2240 pounds
1 gram (gm) = 100 milligrams (mg)	1 pound (lb) = 16 ounces (oz)

Conversion Factors for Mass

1 kilogram = 2·204 pounds
1 pound = 453·6 grams

8. Unit of Time

The **mean solar second** is the unit of time in both systems. It is one 86400th part of a *mean solar day*. The *solar day* is the interval of time between two successive transits of the sun across the meridian at any place. Because the earth's orbit round the sun is elliptic and the earth moves with a non-uniform speed in its orbit, solar days vary in length. The average value of the solar day throughout the year is taken as the mean solar day.

1 mean solar day $=24 \times 60 \times 60=86{,}400$ mean solar seconds.

9. Units of Volume or Capacity

The unit of volume is the volume of a cube of unit length. In the C.G.S. system the volume of a cube of edge 1 cm. is 1 **cubic centimetre**, which is the unit of volume. Another unit, called the litre, is also frequently used for measuring capacity. The **litre** is the volume of one kilogram of pure water at its maximum density, i.e., at 4°C. It is very slightly greater than 1000 cubic centimetres. For most practical purposes the *millilitre*, which is one thousandth part of one litre, may be taken to be equal to 1 cubic centimetre. A cubic centimetre is written as c.c. or cm^3 in abbreviation.

In the F.P.S. system volume is measured in *cubic feet*. Volumes of liquids are generally measured in 'gallons'. A **gallon** is the volume of 10 pounds of distilled water at 62°F.

10. Measurement of Length

In measuring a length you may be called upon to use a scale, such as the *metre scale*, which is graduated in centimetres and millimetres. Or, it may be a foot-rule in which the divisions are in inches and fractions thereof.

In using a scale, first make sure of the unit (inch or centimetre) in which it is graduated and also the value of the smallest scale division. Take the following precautions.

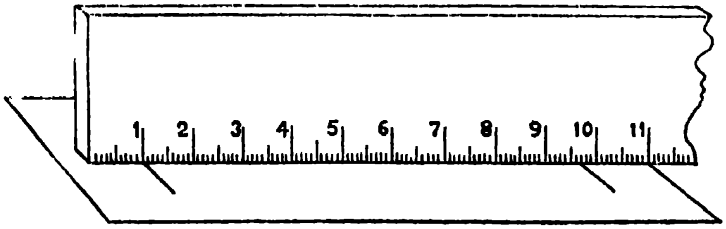

Fig. 2

(i) Place the ruler on its side (as shown in Fig. 2) so as to bring the division marks of the scale as close as possible to the points the distance between which you want to measure.

(ii) Keep your eye vertically over the point at which you are taking the reading.

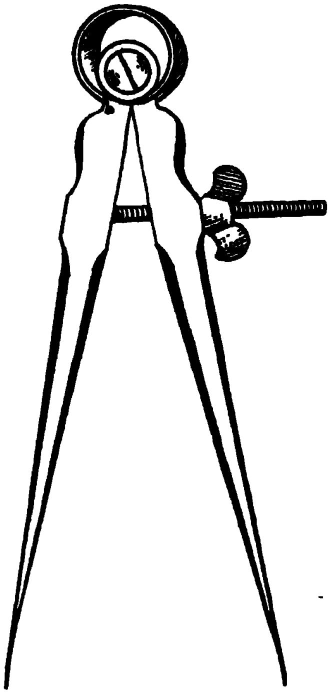

Fig. 3

(iii) Measure the distance by difference, as illustrated in Fig. 2.

When it is inconvenient to place a scale close enough to a body for measuring a distance on the body, you may use a pair of dividers (Fig. 3). The dividers are set to the length to be measured and then transferred to the scale. Those provided with screws (Fig. 3) for setting to small intervals are very useful in workshop practice.

Experiment 1. *To measure the diameter of a sphere with a scale.* Place the sphere between two parallel wooden blocks (Fig. 4). Measure with a scale the distance between the faces of the blocks. Repeat the observation thrice by turning the sphere and using different parts of the scale. Record as follows:

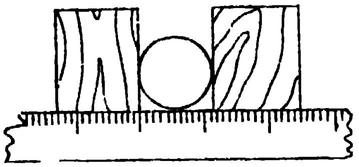

Fig. 4

Diameter of the sphere = (i)....... (ii)........, (iii)..........

Average diameter $= \frac{(i) + (ii) + (iii)}{3}$

11. Simple Calipers

Simple calipers (Fig. 5) may be conveniently used to measure the outside diameter of a sphere or the inside diameter of a tube.

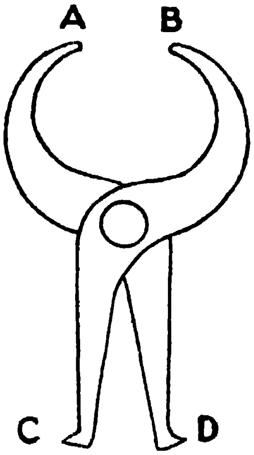

Fig. 5

Experiment 2. *To measure the diameter of a sphere with outside calipers.*

Open the calipers such that the sphere just slips between the jaws AB (Fig. 5) without being scratched. Place the calipers on a scale and read off the distance AB by difference. Take measurements at three different positions of the sphere.

Diameter of the sphere = (i)...., (ii)...., (iii).....

Average diameter $= \frac{(i + ii + iii)}{3} =$ ....

Experiment 3. *To measure the inside diameter of a tube with inside calipers.*

Insert the lower end of the calipers into the tube and stretch the jaws *CD* (Fig. 5) such that they touch lightly the inner surface of the tube. Read off the distance from a scale.

Inside diameter of the tube=(i)....,(ii)....,(iii)......

$$\text{Average diameter}=\frac{(\text{i}+\text{ii}+\text{iii})}{3}=\ldots\ldots\ldots\ldots\ldots$$

12. Eye-estimation

In taking a reading you will often find that it does not coincide with any of the division marks of the scale, but lies between two. A simple way of taking this fractional part of the division into account is to make an estimate of it with the eyes: It is not difficult to guess if the extra length is half a scale division. With practice you can make this estimate more accurate.

There are, however, other and better ways of measuring such fractions. The *diagonal scale* and the *vernier* are the commonest of them. These are described below.

13. The Diagonal Scale

The scale may be graduated in inches or in centimetres and is capable of measuring a length upto 0·01 inch or 0·01 cm. It is divided into ten equal spaces parallel to its length (Fig. 6). The inch (or centimetre) division to the extreme right is further subdivided by ten equidistant diagonal lines.

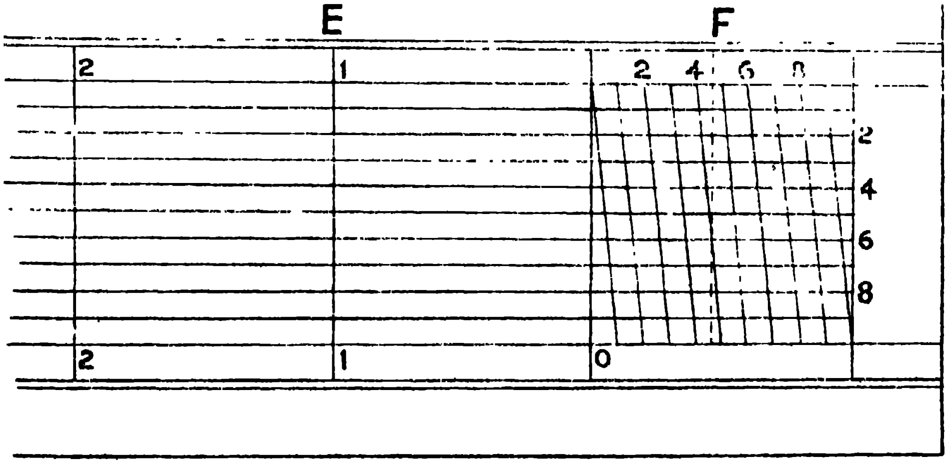

Fig. 6

To measure the distance between two points with a diagonal scale, take a pair of *dividers* and open the jaws so as to make the ends agree with the points. Transfer the divider to the scale.

Suppose the length to be measured is between 1 and 2 inches. Place one leg of the divider on the 1-inch division (*E* ; Fig. 6). Let the other leg be at *F*. It is easy to see from the graduations that the required length is between 1·4 and 1·5 inches.

To find the next place of decimal, move the divider across the scale downwards and parallel to itself, one end always remaining on the vertical line through *E*. The other end will describe the dotted line as drawn in Fig. 6, through *F*. This line cuts only one of the diagonal lines (in this case the fourth) at the point where the latter is cut by the horizontal line sixth from the top. It may be readily seen from this that the length *EF* is 1·46 inches. The first decimal place is obtained from the figures at the top of the right-hand square, and the second decimal place, from those at the right.

Note that the zero of the scale starts from the left of the divided square.

The scale may be used with profit for measuring distances on charts and drawings.

14. The Vernier

The *vernier* is an auxiliary scale which is used in conjunction with a common scale, called the main scale, for measuring fractions of the smallest division of the main scale.

To understand and use a vernier, first find the value of the smallest division on the main scale. Suppose it is 1 mm. Bring the zero of the vernier scale into coincidence with any main scale division. If you look to the other end of the vernier you will find that its last division mark coincides with another division mark of the main scale. Suppose in your case 10 divisions of the vernier scale are equal to 9 divisions of the main scale, i.e., to 9 mm.

10 divisions of the vernier = 9 mm.
1 division „ „ „ = ·9 mm.

The difference between the smallest main scale division and a vernier division is known as **vernier constant**. Therefore, in the above case,

Vernier constant = 1 M.S. division − 1 Vernier division
= 1 mm − ·9 mm = ·1 mm = ·01 cm.

Whenever you want to use a vernier scale, first find the vernier constant.

To measure the length of a rod with a vernier, place one end of the rod against the zero of the main scale and bring the zero of the vernier against the other end. It is clear that the position of the zero of the vernier on the main scale gives the required length. Suppose the position is as in Fig. 7. If you look carefully you will see that the reading is 5·3 (main scale) + a fraction. This fraction is the distance between the 5·3 mark on the main scale and the zero of the vernier. Looking to the divisions of the vernier, suppose, you find that the 7th division of the vernier coincides with a division mark of the main scale. Note that this is also the 7th division mark from 5·3. Therefore, the fractional reading

$$\begin{aligned} &= 7 \text{ M.S. divisions} - 7 \text{ Vernier divisions} \\ &= 7(1 \text{ M.S. div.} - 1 \text{ V.div.}) \\ &= 7 \times \text{Vernier constant} \\ &= 7 \times {\cdot}01 = 0{\cdot}07 \text{ cm} \end{aligned}$$

The length of the rod $= 5{\cdot}3 + {\cdot}07 = 5{\cdot}37$ cm.

To get the reading corresponding to the zero of the vernier, read the main scale first. Then look for that vernier division which coincides with a division mark of the main scale. Multiply the vernier reading by the vernier constant and add to the main scale reading. In general, if the xth division of the vernier coincides with a division mark of the main scale, the vernier reading is $x \times$ vernier constant.

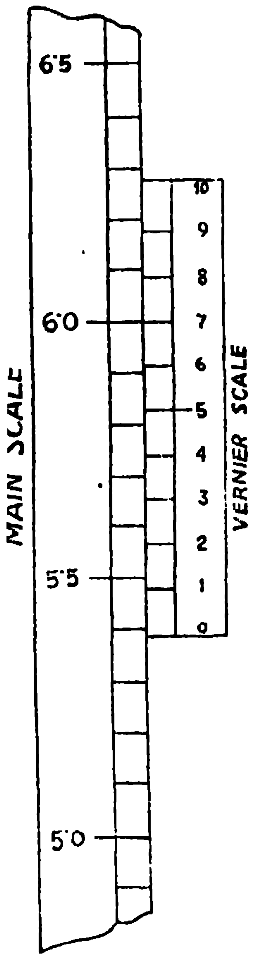

Fig. 7

Verniers may be constructed to read other fractions besides tenths. It may be $\frac{1}{20}$, $\frac{1}{25}$, $\frac{1}{50}$ or any other convenient fraction. In general, if it is to be $\frac{1}{n}$. $n - 1$ divisions of the main scale are made into n divisions of the vernier.

Circular verniers. Verniers are also used to measure angles accurately such as in spectrometers, sextants and theodolites. The divisions are in degrees and minutes of arc.

15. Vernier Slide Calipers

The instrument is an application of the principle of the vernier described above. If you look at Fig. 8, you will notice that it has two jaws AB and CD. The body whose length has to be measured is held between the jaws AB. The jaws CD are used to measure the inside diameter of a tube or hole, and the tail E, for depth. If the instrument is not defective, the zero of the vernier should coincide with the zero of the main scale when the

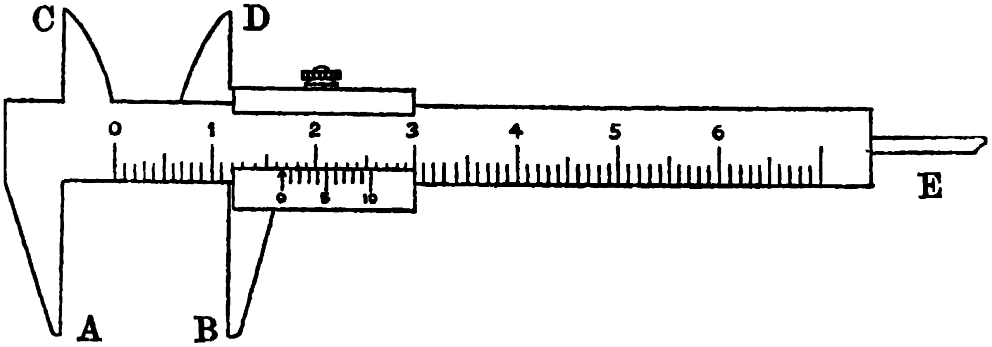

Fig. 8

jaws *AB* are in contact. If the two zeros do not agree, there is a *zero error*. Always examine your instrument for zero error before you use it. See whether the zero of the movable scale is to the right or left of the zero of the fixed scale and find by how much. If to the right, the zero error is to be subtracted from the final reading. If it is to the left it should be added.

Experiment 4. *To measure the thickness of a wooden block.*

Hold the block lightly between the jaws *AB*. Note the readings as follows :

Vernier constant=........cm.
Zero error=........cm.

No. of readings	*Main Scale reading*	*Vernier reading*	*Total*	*Average*
1				
2				
3				

The corrected average thickness =........cm.

Experiment 5. *To measure the inner diameter of a metal washer.*

Insert the jaws *CD* into the opening to touch the sides lightly.

Vernier constant =........cm.
Zero error =........cm.

No. of readings	*Main Scale reading*	*Vernier reading*	*Total*	*Average*
1				
2				
3				

The corrected inside diameter of the washer=......cm.

Experiment 6. *To measure the depth of a hole.*

Insert the tail *E* into the hole such that the end of the main scale rests on the top of the hole. Take readings as follows :

Vernier constant=........cm.
Zero error=........cm.

No. of readings	*Main Scale reading*	*Vernier reading*	*Total*	*Average*
1				
2				
3				

The corrected average depth of the hole=....cm.

16. The Screw Gauge

It is an instrument by which small lengths can be measured with great accuracy, e.g. to ·001 in. or ·001 cm. It consists of a U-shaped piece of steel having two arms. A fixed stud is attached to one arm and the other arm is provided with a cylindrical tube on which a linear scale, called the *fixed scale,* is etched. An accurately cut screw, provided with a collar, works in the tube and moves axially when it is rotated by the milled head. The bevelled end of the collar is divided into a number of equal divisions.

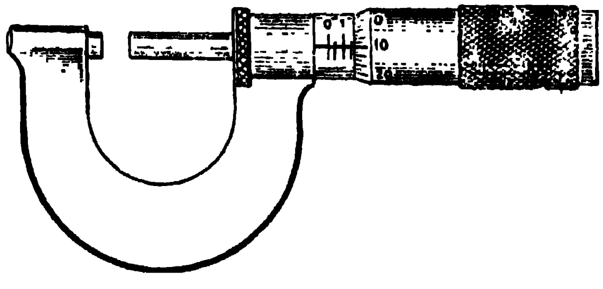

Fig. 9

When the screw is given a complete turn, its end moves forward by the distance between two successive threads of the screw. This is called the *pitch* of the instrument. The axial displacement of the screw for a rotation of one circular division is called the *least count.* If the pitch of the screw is l mm, and there are n circular divisions on the collar, then the least count= l/n mm.

To determine the pitch, give the screw head a number of complete turns. From the change in the linear scale reading you can find the pitch. Note the total number of divisions on the circular scale. Divide the pitch by the number of circular divisions. This gives the least count of the instrument.

To find the zero error of the instrument, the screw head is so turned that the end of the screw just touches the fixed stud.

If in this position the zero mark on the circular scale coincides with the reference line across the fixed scale, there is no zero error. If the zero crosses the reference line, the zero error has to be added to the apparent reading. But if it fails to reach the line, subtract the error from the apparent reading.

Experiment 7. *To measure the diameter of a wire and the area of its cross-section.*

Find the pitch, least count and zero error. Hold the wire lightly between the fixed and the movable studs. **Don't press hard.**

Take the fixed scale reading and note the division of the circular scale which coincides with the reference line. Repeat three times, rotating the wire.

Results : Pitch of the instrument$=$......mm.

No. of divisions on the circular scale$=n$.

Least count of the instrument$=\frac{\text{Pitch}}{n}=$....mm.

Zero error $=$........mm

(Note whether to be added or subtracted).

No. of readings	*Linear scale reading*	*Circular scale reading*	*Total*	*Average*
1				
2				
3				

The corrected average diameter$=$....mm$=$....cm.

Cross sectional area of the wire :

If the diameter of the wire is$=d$ cm. the area of cross section$=\frac{\pi d^2}{4}$ sq. cm.

17. The Spherometer—Measurement of Radius of Curvature

The spherometer is an instrument specially designed for measuring the radii of curvature of spherical surfaces. It consists of a micrometer screw moving through a nut supported on three legs (Fig. 10). The points A, B and C of the legs form an equilateral triangle. The point P of the screw passes through the centre of the triangle ABC. The circumference of the screw head is divided into 100 equal divisions. The pitch of the screw is ordinarily 1 mm, so that each division corresponds to ·01 mm. A linear scale gives the number of complete revolutions which the head makes.

To set the instrument it is placed on a sheet of glass and the screw turned until its tip just touches the plane. This position may be determined by looking tangentially along the glass and finding when the gap between P and its image in the glass just vanishes. If the screw is pushed beyond this position, the instrument rocks. Note the reading of the linear scale division just below the screw head, and also the reading of the circular scale division just against the linear scale.

Now place the instrument on the spherical surface whose radius of curvature is to be measured. Raise or lower the screw until its tip touches the curved surface. Read the linear and circular scales as before. Let the difference between this reading and that on the plane glass be h.

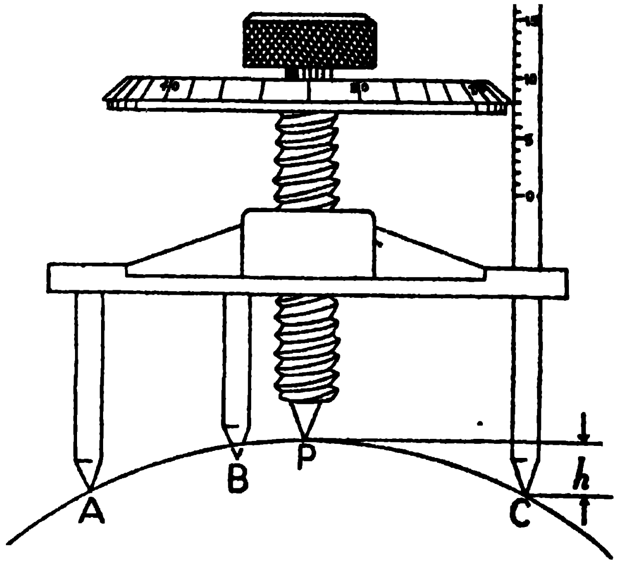

Fig. 10

Screw up the central leg, place the instrument on a stiff piece of paper and press it lightly so that an imprint of the three legs is made on the paper. Measure each side of the triangle formed by the three points and take the mean. Let it be L. Then the radius of curvature 'r' of the spherical surface is given by

$$r=\frac{L^2}{6h}+\frac{h}{2}$$

Enter your results as follows :

Experimental Results

Value of the smallest division on the linear scale of the spherometer=..........cm.

Pitch of the instrument=......cm.

Number of divisions on the circular scale=

Least count=......cm.

Readings on the glass plate :

Number of observations	*Linear scale*	*Circular scale*	*Total*	*Mean* (h_1)
1				
2				
3				

Reading on the spherical surface :

Number of observations	*Linear scale*	*Circular scale*	*Total*	*Mean* (h_2)
1				
2				
3				

$\therefore\ h = h_1 \backsim h_2 = \ldots\ldots\ldots\ldots$ cm.

Reading for L :

(i)......cm, (ii)......cm, (iii)......cm.

Mean $L = \ldots\ldots\ldots\ldots$ cm.

$$\therefore\ r = \frac{L^2}{6h} + \frac{h}{2} = \ldots\ldots\ldots\ldots \text{cm.}$$

18. Measurement of an Area

It is easy to work out from simple rules and formulas the areas of regular figures, such as triangles, rectangles, circles etc.

Areas of irregular figures do not lend themselves to such treatment. They can be found graphically using squared paper.

Experiment 8. The boundary of a farmland, of which the area is to be determined, is shown in Fig. 11. Suppose the scale of the drawing is 1 in. = 1 mile. A piece of graph tracing paper (small division, 0·1 in. and area of small squares, ·01 sq. in.) is placed on the drawing and its outline is marked on the tracing paper.

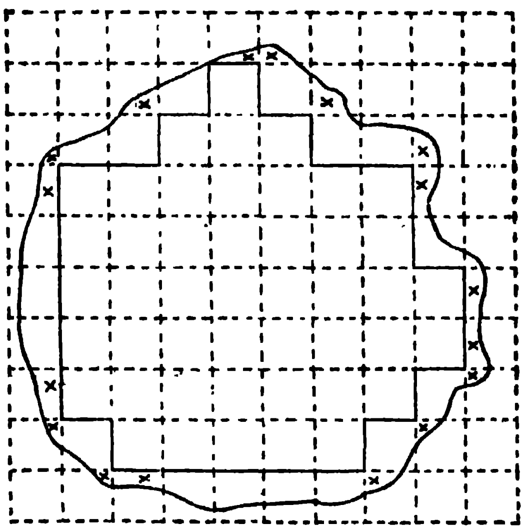

Fig. 11

Count the number of complete squares (·01 sq. in. areas) wholly included within the boundary. Where the boundary line of the area cuts a small square, count it as one square when more than half the area is included within the figure. If less than half the area is included, neglect the square altogether.

If the total number of small squares so obtained is n, the area is $n \times \cdot 01$ sq. mile, since one small square represents ·01 sq. mile,

19. Accuracy in Measurement and in recording Data

We have said that measurement is the foundation on which the exact sciences like physics and chemistry are built. Measurements are however of little value unless carried out with the *proper accuracy*. 'Proper' accuracy does not necessarily mean a high accuracy. You would be satisfied if you knew the distance of your home from the nearest railway station in miles and furlongs. (A furlong is an eighth of a mile). You will not bother about feet and inches. When you ask your carpenter to make you a table 2 ft. 6 inches high, you will not allow him to deviate even by one inch. When you want to replace a portion of the water pipe in your house, you will not allow the diameter of the pipe you purchase to differ from the old pipe by even as much as a fraction of an inch.

These examples serve to show that our requirement of accuracy is not everywhere the same. *The smaller the quantity to be measured the finer has the measurement to be.* This would also involve use of finer measuring instruments.

Accuracy in recording data

In recording data of measurements always use *decimals*, and not fractions. But note that while 1·2 and 1·20 mean the same thing mathematically, they do not do so when they refer to measurements. When we speak of a length as 1·2 cm. we imply that the value lies between 1·1 and 1·3 cm and that the last figure of the measure (i.e., the figure 2 in 1·2) is approximate to this extent. The measurement has not been carried beyond the first place of decimal. When we say that a length is 1·20 cm, we have carried the measurement to the second place of decimal and know that the value lies between 1·19 and 1·21. The zero would show that the latter measurement has been carried out with a greater precision than the first. In physical measurements, a zero at the end of a decimal figure, therefore, carries a definite significance. By failing to record it when it is known to be there, you sacrifice your claim to the accuracy with which you have performed the measurement. On the other hand, if you put it there without making sure that it should be there, you pretend an accuracy to which you have no claim.

This leads us to a question which often puzzles the student. To how many digits should he calculate the final result ? As an exercise you can perform a simple experiment as you read it. With a sharp pencil and a scale, mark on a piece of paper two points exactly 7 inches apart. Then measure the distance

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between them with a scale graduated in centimetres and millimetres. It is likely that you will find a value lying between 17·7 and 17·9 cm. From your data find how many centimetres there are to the inch. How far will you push the division by 7? To two digits, three digits, or more?

Here you do not know the value of the distance in inches probably beyond the first decimal place. It is 7·0 inches. You do not know the second decimal place. This datum is known only to two significant digits. The claim of accuracy in the calculated value cannot go beyond two significant digits. From your measurements you can claim that there are 2·5 centimetres to the inch; but any claim to the next decimal place will be without basis. To get the final value to three significant digits, all data must be known to as many digits.

In calculating the volume of a rectangular slab you may find that the value is 4·23 cm × 2·67 cm × 1·52 cm. On carrying out the multiplication the volume works out to be 17·167032 c.c. Another student may find the size to be 4·24 cm × 2·65 cm × 1·54 cm which on multiplication gives 17·303440 c.c. How should each record his results? We find that in all the measured values the third significant digit is approximate. (4·23 cm means that the value lies between 4·22 and 4·24 cm etc.) Hence the final result need not be given beyond the third significant digit. The recorded final values will be 17·2 c.c. for the first and 17·3 c.c. for the second.

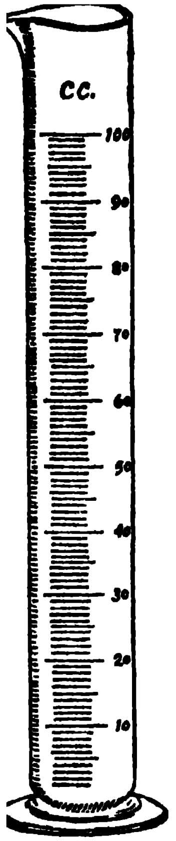


Fig. 12

20. Measurement of Volume

The volume of an object is the amount of space it occupies.

(i) **Volume of a regular solid.** If the object has a regular geometrical shape, its volume can be found by measuring its linear dimensions.

Volume of a rectangular block = length × breadth × thickness.

Volume of a parallelopiped (a solid of which the opposite faces are parallel; each face a parallelogram) = Area of one face × perpendicular distance between this and its opposite face.

Volume of a cylinder $= \frac{\pi}{4} \times \text{length} \times (\text{diameter})^2$.

Volume of a sphere $= \frac{\pi}{6} \times (\text{diameter})^3$.

Volume of a right circular cone $=\frac{1}{3}\times$ area of base $\times$ altitude.

(ii) **Volume of a liquid.** This may be found by pouring the liquid into a graduated *measuring cylinder* (Fig. 12). The surface of a liquid in a narrow vessel is not flat, but curved. To find the position of the liquid surface in the measuring cylinder, you must look with your eye at the same horizontal level as the central part of the surface (Fig. 13). A piece of white paper held behind the vessel will help you in getting the reading accurately.

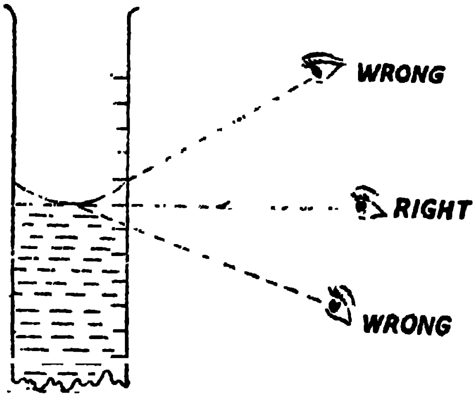


Fig. 13

Measuring cylinders are generally graduated in cubic centimetres. Make sure of the value of each graduation before taking a reading.

A *burette* is a long, narrow, graduated cylinder (Fig. 14). It can measure out a volume of liquid with a higher accuracy than the measuring cylinder.

(iii) Volume of an irregular solid

(a) Put some water in a measuring cylinder and read off the volume. Slide the solid into the cylinder so as not to break the cylinder nor splash out any liquid. Read the volume again. If the solid is fully immersed in the liquid the difference of the two readings gives its volume. (Will the method apply if the solid is soluble in water?)

(b) Get a wide vessel with a spout (Fig. 15). Fill it with water till it overflows. When overflow has stopped place a beaker below the spout to catch any fresh overflow. Now slowly immerse the solid in the water. The solid will displace its own volume of the water, which will come out of the spout. Measure the volume of the water which so comes out. This is also the volume of the solid.

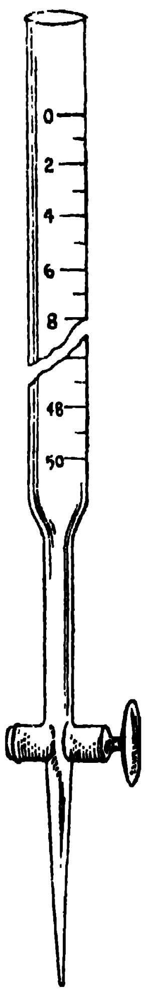


Fig. 14

21. Experiments on the Pendulum

A pendulum suitable for the experiments we want to perform here may be made by tying a thin cotton thread to a round metal bob about 100 gm or so in weight (Fig. 16). The thread may be about a metre long. Its top end may be fastened between two pieces of wood, or it may be held between the jaws of pliers fixed to a stand. A complete oscillation of the pendulum means a full swing, i.e., motion from one extreme end to the other and back again. The time required for a complete oscillation is called the **periodic time**. The maximum displacement that a pendulum undergoes in the course of its swing from its normal position of rest is called its *amplitude*.

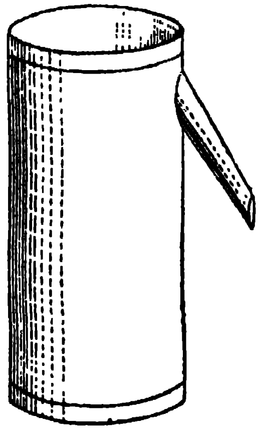


Fig. 15

The periodic time can be measured with a stop-clock (Fig. 17) or a stop watch (Fig. 17-A). These are clocks or watches which can be started or stopped at will by operating a lever or button.

Experiment 9. *To show that the periodic time of a pendulum is independent of the amplitude.*

Apparatus : Pendulum about a metre long, stop-clock, metre scale.

Method : Allow the pendulum to oscillate with an amplitude of, say, 3 cm. Measure the time required for 20 oscillations. Repeat the observations for amplitudes of 4, 5, 6, 7, and 8 cm. You will find that the times for 20 swings will be the same.

Dividing the time by the number of swings, you get the periodic time. *Start counting time when the pendulum is passing through its mean position,* i.e., the position of permanent rest.

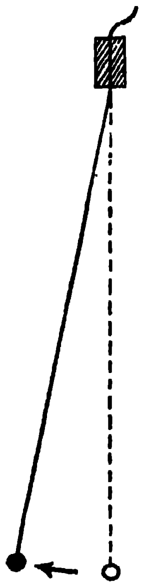


Fig. 16

(Instead of 20 oscillations you may take the time for 25 or 30 oscillations.)

The table shows that the periodic time is the same whatever the amplitude *provided it is small.* This means that the oscillations are executed in equal times which do not depend on the amplitude. This result is often expressed by saying that the oscillations of a pendulum are 'isochronous'.

Results :

Approximate amplitude	*Time for 20 oscillations*	*Periodic time*
3 cm.	... sec.	... sec.
4 cm.	... sec.	... sec.
...	...	...

Experiment 10. *To show that periodic time does not depend on the mass or material of the bob so long as the length is constant.*

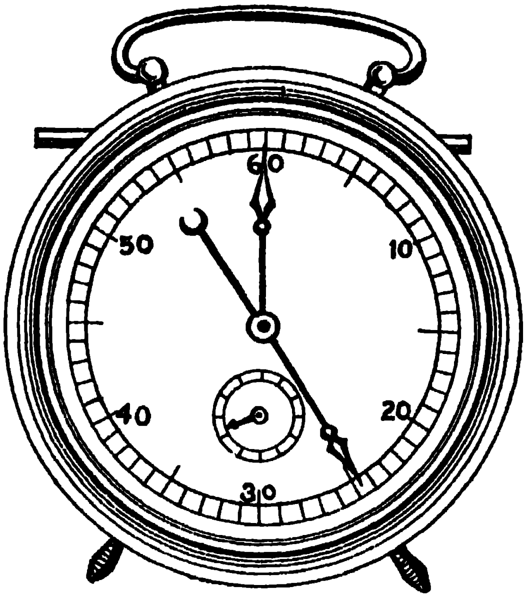


Fig. 17

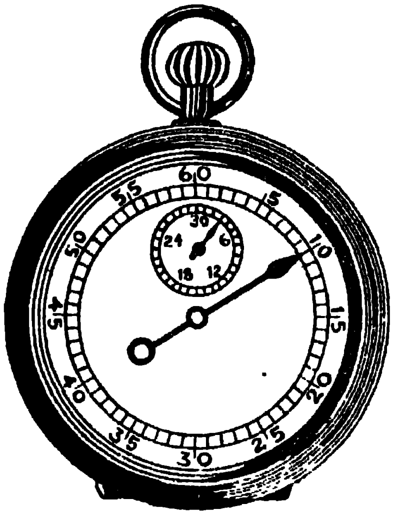


Fig. 17-A

Note that the length of a pendulum is taken to be the distance between the point of suspension and the centre of the spherical bob.

Apparatus : Pendulum bobs of the same size but of different materials, cotton thread, stop-clock, metre scale, calipers.

Method : Using one of the bobs measure the periodic time for 20 oscillations as in the previous experiment. Replace the bob by another, but make sure that the distance from the point of suspension to the centre of the bob (or its bottom, if the bobs are of the same size) remains the same, and repeat. Record as follows:

Material of bob	Diameter of bob	Length of cotton	Total length of pendulum	Time for 20 oscillations	Periodic time
Iron	d cm	l cm	$l+d/2$	... sec	... sec
...	...	...	...	...	...

(If values in column 4 are the same, those in columns 5 and 6 will remain constant whatever we have in columns 1 and 2.)

Experiment 11. *To show that the periodic time is proportional to the square root of the length of the pendulum.*

Apparatus : Pendulum, stop-clock, metre scale, calipers.

Method : Measure time for 20 swings of small amplitude. The lengths may be adjusted to have the values 144, 121, 100 and 81 cm. Remember that the length of the pendulum is the distance from the point of suspension to the centre of the bob. This would require measuring the diameter of the bob.

Results : Diameter of bob=......cm.

Effective length of pendulum, l, in cm.	$\sqrt{l}$	*Time for 20 swings in sec.*	*Periodic time T in sec.*	$\frac{\sqrt{l}}{T}$
144	12	...	...	...
121	11	...	...	...
100	10	...	...	...
81	9	...	...	...

Values in the last column will be the same within experimental error. This shows T is proportional to $\sqrt{l}$.

It will be profitable to draw a graph plotting $\sqrt{l}$ along the x-axis and T along the y-axis. Within experimental error, *the graph will be a straight line passing through the origin.*

22. Mass and Weight

In everyday life we buy and sell many things by '*weight*'. *The weight of a body is the force with which the earth pulls it.* But in purchasing a pound of sugar or a kilogram of meat we are not interested in the force with which the earth pulls it. What we are interested in is the amount we get. *The amount of matter contained in a body is called its mass.*

The distinction between a 'mass of one pound' and a 'force of one pound' or a 'weight of one pound' should be clearly understood. A force (or weight) of one pound is the amount of the earth's pull at its surface upon a mass of one pound. Unfortunately, we often fail to distinguish between the two and use the same word *pound* to mean sometimes an amount of matter, and at other times the pull of the earth upon this amount of matter. The two ideas are clearly different, for *the amount of matter in a body is the same wherever on earth it is situated, while the pull of the earth upon that amount of matter decreases as we recede from the earth's surface.* At any place the weight of a body is proportional to its mass.

Confusion may be avoided if we use the term 'pound' to denote exclusively an amount of matter, and the term 'pound weight' to mean the pull of the earth upon the mass. Similarly for the term 'gram' and 'gram weight', 'ton' and 'ton weight' etc. But writers are hardly consistent in this respect, and it is not infrequent to find a term like 'pound' being used in the sense of a force of one pound, particularly when there is no doubt as to whether a mass or a force is in question.

23. The Balance

To determine the mass or weight of a body in the laboratory, we may use a *spring balance* or a *common balance*.

Spring Balance. It consists of a helical spring (Fig. 18) the upper end of which is fixed to the body of the instrument. The lower end is attached to a rod ending in a hook. The body to be weighed is suspended by this hook. A pointer attached to the rod moves over a scale graduated in grams or pounds. The extension of the spring is proportional to the weight of the body suspended. The position of the pointer gives the mass or weight of the body.

The instrument is quick and convenient to work with. It is capable of weighing a body correct to about 1%. For much of the early work in a laboratory a spring balance may be used. This will save time and also damage to more delicate balances at the hands of beginners. Ranges of 100 gm, 250 gm and 500 gm will ordinarily suffice.

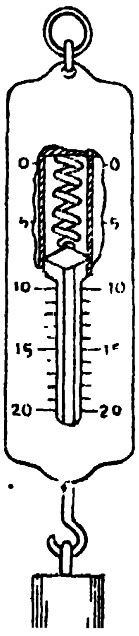


Fig. 18

Common Balance. The essential parts of a common balance are :

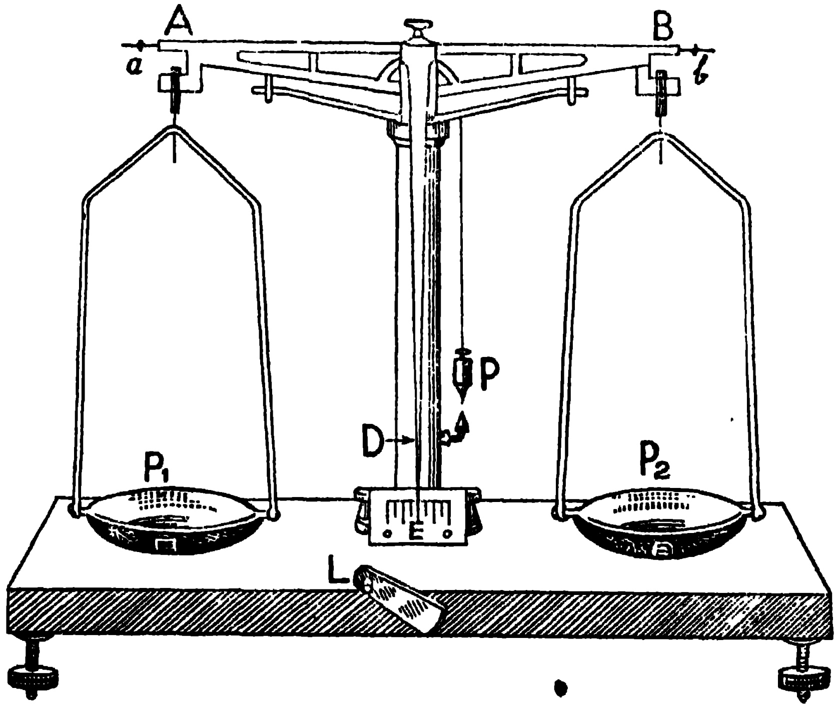


Fig. 19

(i) The balance beam (AB of Fig. 19). It has at its middle a steel knife edge with the edge turned downwards and resting on an agate plate. The beam can oscillate about the knife edge. (The knife edge cannot be shown in the diagram as it is hidden from view by other structures. We shall, however, refer to its position as *C*.) *CA* and *CB* are called the *arms* of the balance. (The balance beam constitutes a lever with equal arms and having its fulcrum at *C*.)

(ii) Two scale pans (P_1 and P_2) supported on stirrups which hang on smaller knife edges near the ends of the beam.

(iii) A vertical pointer *D* attached to the beam near the knife edge *C*. The lower end of the pointer moves over a horizontal scale *E*.

(iv) Two screws, *a* and *b*, at the ends of the beam. They are used to bring the pointer to the zero of the scale.

The balance beam and the pans normally rest on supports and cannot oscillate while in that position. The body to be weighed is placed on the left pan (unless one is left-handed) and the standard weights on the right while the beam is resting on its support. When freed from the support by turning the lever *L*, the beam oscillates.

The instrument is enclosed in a glass case in order that air-currents may not disturb the oscillations of the beam. The balance box is supported on three levelling screws which help to keep the supporting rod of the beam vertical. This adjustment may be checked by a plumb-line (*P*) with which the balance is provided.

The *standard weights* ordinarily contained in the *weight boxes* have values 200, 100, 50, 20, 20, 10, 5, 2, 2 and 1 gram. These are generally made of brass. There are also smaller weights, made of aluminium, having values 500, 200, 200, 100, 50, 20, 20 and 10 milligrams. *Do not handle the weights with the fingers*; *use the forceps provided in the weight box.*

To weigh a body, you must handle the weights in a regular order. Suppose you expect the body to weigh less than 200 grams but more than 100 grams. Try the 200 gm weight first. If it is too high remove it and try the 100 gm. If this is too low, keep it on the pan and add the 50 gm. In this way you go down to the lower weights until you obtain a balance. Count the total and write down the result before you remove the weights. You may find it wiser to record each weight as follows, and finally check before it is put into the weight box.

Weight of body$=(100+50+10+2+1)$ gm $+$ $(200+50+10)$ mgm$=163{\cdot}26$ gm.

Exercises

1. Define the fundamental units in the metric system. Why are they called fundamental ?

What are the advantages of the metric system over the British system ?

2. Define the litre and the gallon. Express the gallon in litres. (Take necessary data from the text to solve the problem).

(*Ans* : 4·55)

Hint : Volume of 62·4 lb of water =1 cft.
∴ ,, 10 lb. ,, ,, =..cft. =..cu.in.
= .... c.c. = ... litre.

3. Express (i) the ton in kilograms and (ii) the kilometre in miles. (*Ans* : (i) 1016 ; (ii) 0·621.)

Hint. (i) 1 ton=....lb. ; 1 lb.=..gm ; ...gm=..kg.
(ii) 1 km=....m ; 1 m.=..in ; 36 in=..yd.

4. Mount Everest is 29141 ft. in height. Express it in metres.

(*Ans* : 8882.2)

5. A 100 yard track has to be extended to 100 metres. Find in feet and inches the elongation necessary. (*Ans* : 28 ft. 1 in.)

6. Express a seer in kilograms, given 1 maund=40 seers =82·2 lb. How much would be a tola in grams if 1 seer=80 tolas ?

(*Ans* : 0·932 kg. ; 11·65 gm.)

7. How much is a mile in kilometres ? (*Ans* : 1·609 km.)

8. A thousand gallon tank has the form of a cube. What is its depth in metres ? (Take cube root of 277 as 6·52). (*Ans* : 1·66.)

9. Distinguish between mass and weight.
Describe the structure of the common balance.

Practical Exercises

1. Draw a straight line 7 inches in length and measure it in centimetres. Hence find how many centimetres there are to the inch.

2. Measure the diameter of the given sphere. Hence calculate its volume. Also find its volume with the help of an overflow can and a measuring cylinder.

(A cylinder may be substituted for the sphere in one set of experiments.)

3. The given card-board has been cut into an irregular shape. Find its area with the help of a graph paper.

You are also provided with a two-inch square of the same board, and a balance. Weigh both boards and find the area of the piece of irregular shape.

4. Measure the periodic time T of a pendulum of length 144 cm, considering 25 oscillations at a time and a total of 100 complete oscillations.

Represent on a squared paper the value of the periodic time T so obtained against the square root of length, $\sqrt{l}$. (Plot $\sqrt{l}$ along the x-axis, 5 small squares representing one unit of $\sqrt{l}$. 25 small squares along the y-axis may represent 1 second of T.)

Draw a straight line joining the point so plotted to the origin. From the graph find the value of l corresponding to a periodic time of 2 seconds.

(A pendulum which takes 1 second to execute a half swing is called a *seconds pendulum*. You have found its length).

Oral Questions

1. What is a *vernier* ? What do you understand by *vernier constant* ? If the vernier constant of an instrument is one-tenth of a millimetre, can you measure with this instrument a length with an error less than this value ?

2. Which is capable of greater accuracy—a vernier caliper or a screw-gauge ? How will you check any of these instruments for a zero error ? How do you decide whether the zero error is to be added to or subtracted from the apparent reading ?

The zero error of a screw gauge is ·05 mm. If the length measured is 1 cm could not the zero error be neglected ?

3. You have measured the diameter of a sphere with a vernier calipers which reads up to 0·1 mm, and found that the value is 2·32 cm. To how many significant digits should you calculate the volume ? Why ?

4. What is the difference between the statements (i) a body weighs 23·5 gm, and (ii) it weighs 23·50 gm ?

5. Does the size of a bob affect the length of a pendulum ? Will there be any harm if, instead of being held between jaws, the upper end of a pendulum is tied round a rod ?

6. When would you prefer a spring balance to a common balance ?

CHAPTER 2

DENSITY AND SPECIFIC GRAVITY

24. Density

We often say that iron is 'heavier' than wood. But a log of wood may be heavier than a piece of iron. How can then both statements be correct at the same time ? In saying that iron is 'heavier' than wood, we mean that a block of iron weighs more, i.e., contains more matter, than a block of wood *of the same volume.*

The confusion of using the word 'heavier' in two different ways may be avoided by introducing the idea of 'density' to mean mass in relation to volume. We may then say 'iron is denser than wood', 'lead is denser than iron', 'mercury is denser than water' and 'water is denser than kerosene'. There is no confusion in these statements.

Density of a substance is the mass of unit volume of it. In the c.g.s. system density is expressed in grams per cubic centimetre. In the f.p.s. system density is measured in pounds per cubic foot. Occasionally, however, the mass or the volume may be expressed in other units too.

Gases are very light compared with solids and liquids. A volume of gas does, however, have a weight and a mass, and hence a density too. The density of a gas depends on its temperature and pressure.

Densities of some common substances
(in gm per c.c.)

Substance	*Density*	*Substance*	*Density*
Brass	8·4 to 8·7	Alcohol (ethyl)	·792
Copper	8·93		
Cork	·22 to ·26	Glycerine	1·26
Gold	19·32	Kerosene	·80
Glass (crown)	2·5 to 2·7	Mercury	13·6
Iron	7·0 to 7·9	Milk	1·03
Marble	2·5 to 2·8	Paraffin Oil	·8
Paraffin	·87 to ·93	Sea-water	1·01 to 1·05
Sand	2·63	Turpentine	·87
Sugar	1·59		

25. Specific Gravity

When densities are expressed in different units they cannot be compared without some calculation. If we are told that the density of stone is 156 pounds per cubic foot, of iron 2·8 pounds per cubic inch and of brass 8·4 grams per cubic centimetre, it will take us some time to find by calculation that brass is denser than iron, and iron denser than stone. For purposes like this it is more convenient to know the *specific gravity* of a substance instead of its density.

The specific gravity of a substance is the ratio of the weight (*or mass*) *of the substance to the weight* (*or mass*) *of an equal volume of water.*

$$\text{Specific gravity} = \frac{\text{Weight (or mass) of substance}}{\text{Weight (or mass) of equal volume of water.}}$$

Note that *specific gravity is a ratio*, and is therefore expressed as a pure number, i.e., without any unit attached to it. Density, however, is expressed in units of mass per unit of volume.

It is easy to see that specific gravity (henceforth abbreviated to S.G.) of a substance is also the ratio of its density, d, to that of water, d'. For if V is the volume of the substance

$$\text{S.G.} = \frac{\text{Mass of volume } V \text{ of substance}}{\text{Mass of volume } V \text{ of water}}$$

$$= \frac{V \times \text{Density of substance}}{V \times \text{Density of water}} = \frac{Vd}{Vd'} = \frac{d}{d'}$$

Since, in the c.g.s. system, the density of water is 1 gm per c.c. or 1 kg per litre, the specific gravity of a substance is equal to the number expressing its density in grams per c.c. or kg per litre.

The density of iron is 7·7 gm per c.c. Hence its S.G. is 7·7. If the S.G. of brass is 8·4, its density will be 8·4 gm per c.c. or 8·4 kg per litre. The table at p. 25 also gives the specific gravities of the substances.

In the f.p.s. system the density of water is 62·4 lb. per cft. (pounds per cubic foot). Hence the density of a substance in f.p.s. units (i.e. lb. per cft.) can be found by multiplying its S.G. by 62·4.

26. Effect of Temperature on Density and Specific Gravity

The volume of a body increases as it gets warmer, but its mass does not change. Since density = mass ÷ volume, it is clear that density will diminish as temperature increases. In solids, the change of volume with temperature is small. Hence unless a high accuracy is aimed at, mention of temperature in connexion with the density of a solid will not be necessary. For the same rise in temperature liquids expand much more than solids, and gases much more than liquids. It is, therefore, better to mention the temperature when speaking of the density of a liquid. In gases, the volume is also very easily changed by pressure. So, we must always mention the temperature and pressure of a gas when speaking of its density.

Specific gravity, we have seen earlier, is the ratio of the density of the substance to that of water. Since the density of water changes fairly rapidly with temperature, it is necessary to say at what temperature we should take the water to be in defining specific gravity. Water is densest at 4°C, and its density at that temperature is by definition 1 gm per c.c. In accurate determinations of specific gravity, it has been agreed that the water should be considered to be at 4°C. It is, therefore, correct to say that:

The specific gravity of a substance is the ratio of the weight (or mass) of any volume of the substance to the weight (or mass) of an equal volume of water at 4°C.

27. Determination of Density

Since density of a substance is its mass per unit volume, a determination of density will involve two measurements, viz., (1) the mass of a body made of the substance, and (2) the volume of the body. Mass can be measured by weighing as detailed in § 23. To find the volume we can apply one of the methods described in § 20

Experiment 12. *To find the density of a solid of regular geometrical shape.*

Apparatus : The given solid (a piece of wood in the form of a rectangular parallelopiped, cylinder or sphere) vernier calipers, balance.

Results (for a cylinder) :

Vernier constant of calipers = ... cm.

Zero error of calipers = ... cm (say whether to be added or subtracted).

Diameter of cylinder at three places (two readings at right angles at each place) :

(i) ...... cm, (ii) ...... cm,
(iii) ...... cm, (iv) ...... cm,
(v) ...... cm, (vi) ...... cm.

Mean corrected diameter, d = ... cm.

Length of cylinder :

(i) ...... cm, (ii) ...... cm, (iii) ...... cm.

Mean corrected length, l = ... cm.

Volume, $V = \frac{1}{4}\pi d^2 l = \frac{1}{4} \times \frac{22}{7} \times \quad \times \quad \times \quad = \ldots\ldots$c.c.
Mass, $M = \ldots\ldots$gm.
Density $= M/V = \ldots\ldots$gm per c.c.

Experiment 13. *To find the density of a solid of irregular shape.*

Apparatus : Solid, measuring cylinder, water, balance.

Results :

Mass of solid (by balance), $M = \ldots\ldots$gm.
Reading of measuring cylinder before adding solid, $V_1 = \ldots\ldots$c.c.
Reading of measuring cylinder after adding solid, $V_2 = \ldots\ldots$c.c.
Volume of solid, $V = V_2 - V_1 = \ldots\ldots$c.c.
Density $= M/V = \ldots\ldots$gm per c.c.

Experiment 14. *To find the density of a liquid.*

Apparatus : Liquid (a solution or an oil), burette and stand, clean dry beaker, balance.

Method : Weigh the beaker. Pour the liquid into the burette and allow a little to run out, but not into the beaker. Take the reading of the burette. Now allow a convenient volume of the liquid to run into the beaker. Take the burette reading. Finally weigh the beaker with the liquid.

Results :

Mass of empty beaker = ... gm.
First reading of burette = ...c.c.
Final reading of burette = ...c.c.
∴ Volume, V, of liquid taken = ...c.c.
Mass of beaker + liquid ....gm.
∴ Mass, M, of liquid taken = ...gm.
Density $= M/V = \ldots\ldots\ldots$gm per c.c.

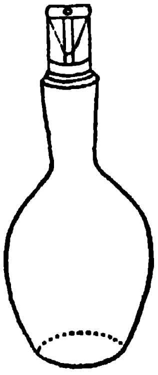


Fig. 20

(ii) The density of a liquid can be accurately determined with the help of a specific gravity bottle (Fig. 20), also known as *density bottle* or *pyknometer*. The bottle is closed with a ground glass stopper, which has a small hole through it to allow liquid to escape when it is put in position. Such bottles are often made to contain a specified volume of water at a specified temperature.

Apparatus : Liquid, clean and dry specific gravity bottle, balance.

Method : Weigh the empty bottle, which must be clean and dry. Fill the bottle with the liquid and insert the stopper. Dry the outside with clean rag, blotting paper or filter paper. There must not be any air bubble inside the bottle. Weigh it now.

Results :

Mass of empty bottle = ... gm.
Mass of bottle filled with liquid = ... gm.
∴ Mass, M, of liquid filling bottle = ... gm.
Inner volume, V, of bottle (as specified) = ... c.c.
Density = M/V = ... gm per c.c.

Bottles of specified volume are rather costly. For the purpose of this experiment any specific gravity bottle, even if its volume is not known, may be used without inconvenience. The bottle, completely filled with water, is weighed and the mass, in grams, of the water completely filling the bottle is determined. Since water has a density of 1 gm. per c.c., the inner volume of the bottle, in cubic centimetres, is easily obtained.

Experiment 15. *To find the density of a powder, or solid in small pieces insoluble in water.*

Apparatus : Sand (or metal shavings), specific gravity bottle, water, balance.

Method : Weigh as indicated below under the head 'Results'. To ensure that there are no air bubbles when you have sand and water in the bottle, add a little water to the sand in the bottle and shake. The quantity of sand may be such as to fill one-third of the bottle.

Results :

Mass of empty bottle, m_1 = ... gm.
Mass of bottle + sand, m_2 = ... gm.
∴ Mass of sand = $m_2 - m_1$ = ... gm.
Mass of bottle + sand + water on sand, m_3 = ... gm.
∴ Mass of water on sand, $m_4 = m_3 - m_2$ = ... gm.
Mass of bottle + water completely filling bottle, m_5 = ... gm
∴ Mass of water completely filling bottle, m_6
$= m_5 - m_1$ = ... gm.
∴ Mass of water having same volume
as sand = $m_6 - m_4$ $(= V)$ = ... gm.
∴ Volume of water having same volume as sand = .. V c.c.
∴ Density of sand = $(m_2 - m_1)/V$ = ... gm per c.c.

Experiment 16. *To find the density of a powder (or solid in small pieces) soluble in water.*

Apparatus : Sugar, specific gravity bottle, water, kerosene, balance.

Method : Proceed as in Experiment 15, but use kerosene instead of water. Sugar does not dissolve in kerosene. In this way find the ratio

$$\frac{\text{mass of sugar} \quad (m_2-m_1)}{\text{mass of an equal volume of kerosene} (V)}$$

(m_1, m_2 and V relate to the notations used in Expt : 15).

This ratio may be called the density of sugar relative to kerosene. If this is multiplied by the density of kerosene, we get the density of sugar. For

$$\text{Density of sugar} = \frac{\text{mass of a quantity of sugar}}{\text{its volume } (V)}$$

$$= \frac{\text{mass of the sugar}}{\text{mass of an equal volume, } V\text{, of kerosene}} \times \frac{\text{mass of volume } V \text{ of kerosene}}{\text{volume } V}$$

=Density of sugar relative to kerosene × density of kerosene.

The density of kerosene may be determined in any one of the ways described in Expt. 14.

28. Determination of Specific Gravity

Determining the specific gravity of a substance does not involve any more measurement than in finding its density. Measuring the mass of the body is common to both. To find the density we must know the volume of the body, whereas to find its specific gravity we must know the mass of an equal volume of water. So the volume also must be known in both cases. Since in the c.g.s. system the density of water is 1 gm per c.c. the mass of the water in grams is the same as its volume in cubic centimetres. To find the specific gravity of a body we shall not, therefore, be required to make any additional measurement beyond those required for finding the density. All methods described in Experiments 12 to 16 for determining density may, therefore, be taken also as methods for measuring specific gravity. After finding the volume V of the solid in c.c., we need only to say that the mass of an equal volume of water $=V$ gm, and hence the specific gravity= M/V, but add no unit as specific gravity is only a number. With the specific gravity bottle, the data directly provide the mass of an equal volume of water.

The Hydrometer. The specific gravity of liquids can be directly determined by means of an instrument known as the **commercial hydrometer.** The commercial constant-weight hydrometer is a hollow glass chamber weighted at the bottom and has a graduated stem at the top (Fig. 21). When floated in a liquid it sinks until it displaces its own weight of the liquid. The lighter the liquid the deeper it settles before coming to rest. The scale attached to the stem is so calibrated that the specific gravity of the liquid can be read directly at the point where the stem projects through the liquid surface.

The commercial value of many liquids such as sugar solutions, sulphuric acid, alcohol etc. depends directly on their specific gravity. Hydrometers are extensively used to find their specific gravities. Some hydrometers are designed and calibrated for specific purposes. An alcoholometer will give the concentration of spirit ; a lactometer will judge the purity of milk. One of the commonest uses of a hydrometer is in testing the acid solution of the lead storage battery.

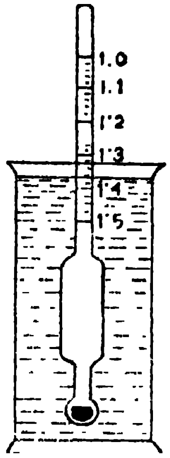


Fig. 21

Exercises

1. Define density. Distinguish between density and specific gravity.

2. Define specific gravity and show that it is equal to the density of the substance divided by the density of water.

3. Why is the specific gravity of a substance equal to the number expressing its density in grams per c.c. ?

4. A lead ball weighs 50 kg. If the sp.gr. of lead is 11·4, what is the volume ?

(*Ans* : 4386 c.c.)

5. The density of sulphuric acid is 1·8 gm per c.c. How much acid is contained in a bottle of volume 2 litres ?

(*Ans* : 3600 gm.)

6. At one time the alloy used for silver coins was 90 parts silver and 10 parts copper by weight. What is the sp.gr. of the alloy ? (Given sp. gr. of silver =10·5 ; of copper=8·8).

(*Ans* : 10·3)

7. An ornament made of a gold-silver alloy has a volume of 30 c.c. and a mass of 525 gm. Find the weight of gold it contains. (Sp. gr. of gold =19·3 and of silver=10·5). (*Ans* : 460·5 gm)

8. A density bottle weighs 24·20 gm when empty, 67·81 gm when filled with turpentine and 74·20 gm when filled with distilled water. Find the density of turpentine.

(*Ans* : ·872 gm per c.c.)

9. Find the density of copper from the following data :
Weight of density bottle empty = 20·6 gm.
Weight of bottle + copper turnings = 111·1 gm.
Weight of bottle + copper + water on top = 125·4 gm.
Weight of bottle full of water = 45·3 gm.

(*Ans* : 8·7 gm per c.c.).

10. Find the density of salt from the following data :
Weight of density bottle empty = 24·20 gm
Weight of bottle + salt = 27·20 gm
Weight of bottle + salt + turpentine on top = 69·91 gm
Weight of bottle full of turpentine = 67·81 gm
Weight of bottle full of water = 74·20 gm

(*Ans* : 2·91 gm per c.c.)

ARCHIMEDES' PRINCIPLE

29. Archimedes' Principle

The story how Archimedes concluded that King Hiero was cheated by a goldsmith, whom the king had asked to make a crown of pure gold, is well known. It was based on the knowledge, believed to have been acquired by him while stepping into a bath tub, that a body immersed in water displaces its own volume. Later on, he also noticed that the body feels lighter when immersed. By careful reasoning he reached the conclusion that an upward force equal to the weight of the displaced liquid acts on the body immersed. This is now known as the Principle of Archimedes. We may state it as follows:

A body wholly (or partly) immersed in a liquid appears to lose a part of its weight, which is equal to the weight of the displaced liquid.

If the weight of the body in air $=W$ gm-wt., and its weight in water $=w$ gm-wt., then the apparent loss in weight$=$ $(W-w)$ gm-wt.

By Archimedes' principle, $(W-w)$ gm-wt. is the weight of the water displaced by the solid. At any place weight is proportional to the mass. It, therefore, follows that if the mass of the body in air $=M$ gm, and its apparent mass in water $=m$ gm, then $(M-m)$ gm is the mass of the displaced water.

Archimedes' principle applies to all liquids and gases. Liquids and gases together are known as 'fluids'. (A fluid means any thing that flows). A more general way of stating Archimedes' principle will be to say

"*When a body is immersed in a fluid, it is acted on by an upward force equal to the weight of the displaced fluid.*" The ascent of balloons is governed by Archimedes' principle.

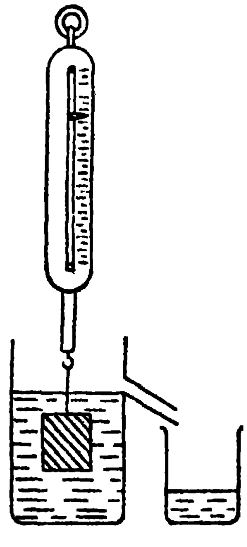


Fig. 22

Experiment 17. *To verify Archimedes' principle.*

(i) *Apparatus* : Solid (a lump of glass, stone or metal), spring balance, overflow can, beaker, measuring cylinder, thread, water (Fig. 22).

Method : Suspend the solid by a piece of thread from a spring balance, and record its weight. Fill the overflow can, and after the overflow has stopped, place an empty beaker under its sidetube. (The first overflow should be caught in a vessel and thrown away). Slowly lower the solid, still suspended from the spring balance, into the overflow can till it is fully immersed. Record the apparent weight of the solid as given by the spring balance. Show that the weight of the water collected in the empty beaker is equal to the difference of the weights of the solid in air and in water.

(ii) *Apparatus* : Solid, spring balance, beaker, water, thread, a metal cylinder exactly fitting into another (Fig. 23).

The cylinder *B* is closed. It fits exactly into *A* and fills *A* completely. The inner volume of *A* is thus equal to the outer volume of *B*. *B* may be withdrawn from *A* and attached below it.

Method : Suspend the cylinders from the spring balance and record the weight, *B* being attached to the bottom of *A* as in Fig. 23. Lower *B* into a beaker of water so as to be fully surrounded by water. Record the apparent weight as given by the spring balance.

Carefully pour water into *A* (from a pipette or a wash-bottle). When it is full note that the initial reading of the balance is restored. This clearly shows that the apparent loss in weight on immersion is equal to the weight of the displaced liquid.

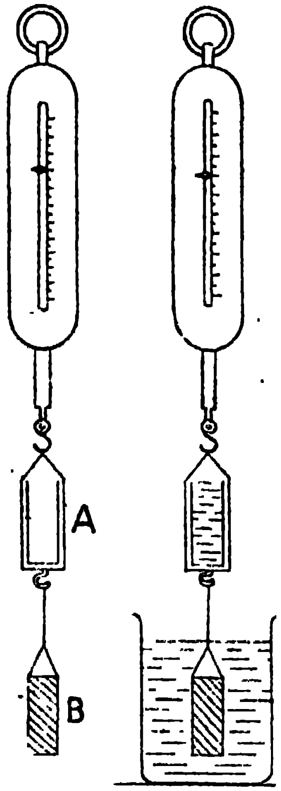


Fig. 23

30. Application of Archimedes' Principle to Determination of Volume

Archimedes' principle states that when a body is weighed (i) in air, and (ii) when *fully immersed* in water, the difference between the observed masses is equal to the mass of the water displaced by the body.

If the mass of the body in air $=M$ gm, its apparent mass when fully immersed in water $=m$ gm,

then $(M-m)$ gm = mass of the water displaced.

Now, 1 gm of water has a volume of 1 c.c,

$\therefore$ Volume of water displaced $=(M-m)$ c.c.
But this is also the volume of the solid.
$\therefore$ Volume of solid $=(M-m)$ c.c.

31. The Hydrostatic Balance

It differs but little from a common balance, and is designed to weigh a body both in air and water with convenience (Fig. 24).

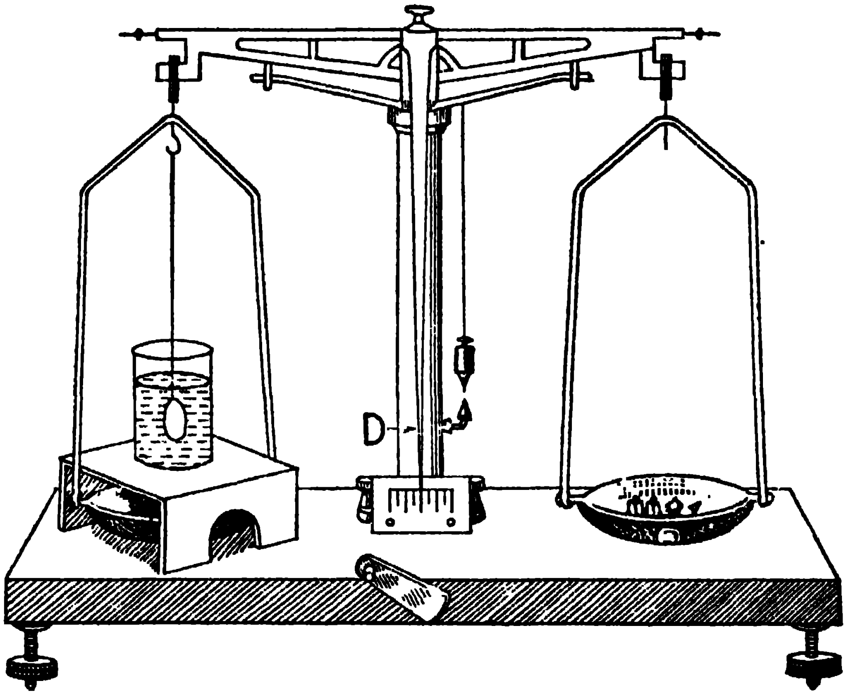


Fig. 24

The balance carries a hook above the pan. The body to be weighed in water is hung by a fine thread from this hook. A bridge spans the scale pan and supports a beaker of water in which the body is immersed. The bridge or the beaker does not touch the pan or its supporting arms. The body is weighed, first, while hanging in air, and then when fully immersed in water. While in water, it must not touch the bottom or the sides of the beaker.

32. Using the Hydrostatic Balance for determining Densities and Specific Gravities

We have just seen how by weighing a solid in air and in water we can find its volume. We have also seen how a hydrostatic balance makes it easy for us to weigh a body both in air and water. This balance, therefore, provides convenient data for finding the density and specific gravity of a solid. Its use can easily be extended to liquids.

In all experiments with the hydrostatic balance it is necessary to take the following *precautions* :

(*a*) *The body must remain completely immersed even when the beam is raised.*

(*b*) *The body must not touch the sides or the bottom of the beaker.*

(*c*) *The bridge must not touch the balance pan.*

(*d*) *No air bubble must stick to the body.*

Experiment 18. *To determine the density of a solid heavier than water.*

Apparatus : A piece of glass (or a metal), thread, hydrostatic balance with weight box, bridge, beaker, water.

Method : Weigh the solid in air and then in water. The thread may be waxed so as not to soak water. Take all precautions mentioned above.

Results :

Mass of solid in air, m_1 =..........gm.
Mass of solid in water, m_2 =..........gm.
∴ Mass of displaced water which has the same volume as the solid $= m_1 - m_2$ =..........gm.
∴ Volume of solid $= (m_1 - m_2)$ c.c. =.......... c.c.

$$\therefore \text{ Density} = \frac{m_1 \text{ gm.}}{(m_1 - m_2) \text{ c.c.}} = \ldots \text{gm. per c.c.}$$

$$\left[\text{S. G.} = \frac{m_1 \text{ gm.}}{(m_1 - m_2) \text{ gm.}} = \ldots\ldots\ldots\ldots \right]$$

Experiment 19. *To determine the density of a solid lighter than water,*

Apparatus : The solid (a lump of wax), sinker (a metal piece), hydrostatic balance, weight box, bridge, beaker, water, thread.

Method : We require another body, called a *sinker*, to be attached to the solid so that the two together may sink in water. First, weigh the solid in air. Tie the sinker to the solid so that you may weigh with the solid in air and the sinker in water without the sinker touching the bottom or the sides of the beaker. Then weigh with both the solid and the sinker fully immersed in water.

Results :

Mass of the solid in air, m_1=..........gm.
Mass of the solid in air + sinker in water, m_2 =....gm.
Mass of solid in water + sinker in water, m_3=....gm.
∴ $m_2 - m_3$ = Mass of solid in air—mass of solid in water
= mass of water displaced by solid, =..........gm.

$\because$ Volume of solid = Volume of displaced water $=(m_2-m_3)$ c.c.
$=$........c.c.

$$\therefore \text{ Density} = \frac{m_1 \text{ gm.}}{(m_2-m_3) \text{ c.c.}} = \text{......gm. per c.c.}$$

$$\left[\text{S. G} = \frac{m_1 \text{ gm.}}{(m_2-m_3) \text{ gm.}} = \text{........}\right]$$

Experiment 20. *To determine the density of a solid soluble in water.*

Apparatus : Solid (large crystal of copper sulphate or alum), hydrostatic balance, weight box, bridge, beaker, water, thread, kerosene.

Method : Using the method described in Experiment 18 first find the density of the solid relative to kerosene.

Multiply the value by the density of kerosene, which may be found from the tables or determined by any of the methods described in Expt. 14 or 21.

Experiment 21. *To determine the density of a liquid.*

Apparatus : Liquid (a solution or an oil), hydrostatic balance, weight box, bridge, beaker, water, thread, a solid which does not dissolve in the liquid or in water, and sinks in both.

Method : Weigh the solid in air, water and the liquid.

Results :

Mass of the solid in air, $m_1=$..........gm.
Mass of the solid in water, $m_2=$..........gm.
Mass of the solid in the liquid, $m_3=$..........gm.
Mass of liquid displaced by the solid $=m_1-m_3=$....gm.
Mass of water displaced by the solid $=m_1-m_2=$......gm.

$\therefore$ Volume of solid $=(m_1-m_2)$ c.c.

Density of liquid

$$= \frac{\text{mass of a volume of liquid equal to the volume of the solid}}{\text{volume of the solid}}$$

$$= \frac{(m_1-m_3) \text{ gm.}}{(m_1-m_2) \text{ c.c.}} = \text{... gm. per c.c.}$$

$$\left[\text{S.G} = \frac{(m_1-m_3) \text{ gm.}}{(m_1-m_2) \text{ gm.}} = \text{..........}\right]$$

Exercises

1. State Archimedes' principle, and describe an experiment to verify it.

2. How can you find the volume of a solid with the help of Archimedes' principle? What precautions should you take so that your result may not be vitiated?

3. A body weighs 50 gm in air and 40 gm in water. Find its volume and specific gravity. (*Ans* : 10 c.c.; 5)

4. A string, which snaps when the load on it is 4 lb., is made to support under water a brick which weighs 5 lb. in air. Find what fraction of the brick will emerge when the string snaps in the course of pulling the brick out of the water. (Sp. gr. of brick =1·5.) (*Ans* : 7/10)

5. A body weighs 300 gm in air and 270 gm in a liquid of sp. gr. 0·9. How much will it weigh in water? Find its volume and specific gravity. (*Ans* : 266·7 gm ; 33·3 c.c. ; 9)

6. The weight of a piece of gold alloyed with silver is 20 gm in air and 18·7 in water. If the specific gravities of gold and silver are 19·3 and 10·5 respectively, how much gold is there in the piece? (*Ans* : 13·9 gm)

7. Find the density of cork from the following data :

Weight of cork in air =2 gm.
Weight of sinker in water =50 gm.
Weight of cork and sinker both under water = 44 gm.

(*Ans* : 0·25 gm per c.c.).

8. A piece of sodium weighs 0·34 gm in an oil of which the specific gravity is 0·8, and 0·54 gm in an oil of sp. gr. 0·7. Find the mass, density and volume of the specimen. (*Ans* : 1·94 gm; ·97 gm per c.c. ; 2 c.c.)

9. Describe and explain how you would determine the density of a coin by weighing it in air and in water.

CHAPTER 4

PRESSURE IN LIQUIDS

33. Pressure—What it means

Hold a pencil vertically with its flat end resting on your palm and load the pencil by putting a book of moderate weight on its top. The weight of the book exerts a thrust on your palm. Now invert the pencil so that its sharp end rests on your palm while the flat end supports the book. The weight of the book exerts the same thrust as before, but the sensation to which it gives rise is more acute and may even be painful if the pencil is very sharp.

What is the reason for this difference in sensation ? The force with which the palm was pressed, was the same in both cases. But the areas over which this force acted were different. With the flat end of the pencil on your palm, the weight of the book was spread over a wider area than when the sharp end pressed you. What mattered was not the force alone, but also the area over which it acted.

It is not difficult to get other examples of the kind. If you wrap round a heavy parcel with a thin string and hold it for some time by the string, you will soon feel pain. But if the string is attached to a wide handle, and the parcel is held by this handle, it could be carried with less discomfort. In the former case the force exerted per unit area on your fingers by the load was larger than in the latter case.

When you walk bare-footed on plane ground the weight of your body acts over that area of your foot which is in contact with the ground. But if you walk on a bed of pebbles, particularly with sharp edges, you feel pain because the weight of your body now acts over much smaller areas.

34. Pressure and Thrust

Pressure is defined as the force per unit area. If a total force F acts over an area A, the pressure P is given by

$$P=\frac{\text{Force}}{\text{Area}}=\frac{F}{A}$$

The force F is often called a *thrust.* It acts perpendicular to the area A and gives rise to the pressure.

Units of pressure. It is clear from the above that 'pressure' is not the same as 'force'. While force is expressed in the usual units, such as pounds or dynes, pressure is expressed in units like 'pounds per square inch' or 'dynes per square centimetre' etc., involving both the unit of force and the unit of area.

In engineering practice, the term 'pressure' or 'total pressure' is frequently used with the meaning of force in the foregoing statement, while 'unit pressure' or 'intensity of pressure' is used to mean 'pressure' as defined above. In this book we shall use the term 'pressure' as force per unit area.

35. Pressure at a Point

We often speak of 'pressure at a point', particularly when discussing liquids. Since a point has no area, a question naturally arises as to the meaning of the term. If F is the force exerted over an area A surrounding the point in question, A being so small that the pressure over it may be taken as uniform, then the ratio $P=F/A$ is called the pressure at the point.

36. Liquid exerts Pressure

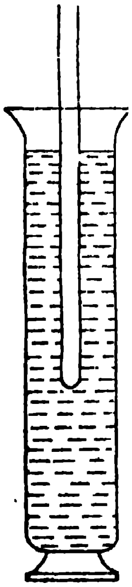


Fig. 25

A liquid exerts a force against any surface with which it is in contact. The existence of such forces may be demonstrated in various ways.

(i) Take a long tube closed at one end and push it down into a column of water as shown in fig. 25. When left to itself the tube will be seen to spring upward. Your pencil, a light stick or a piece of cork will do the same. Obviously, water exerts an upward force upon the bottom of each.

A boat would sink if there were no such force. The force balances the weight of the boat.

(ii) Take a tall cylinder, closed at one end and provided with small holes at different depths below the open top. When you fill it with water you will find that (a) water comes out with greater force at greater depths, and (b) from each hole the jet of water emerges perpendicular to the wall of the cylinder (Fig. 26). (This latter effect will be masked if the jets are made to pass through glass tubes fitted to the side of the cylinder.)

These show that (a) *the liquid exerts a thrust which increases with depth* (below the top), and that (b) *the thrust acts normal to the surface* (of the container).

(iii) *The thrust exerted by a liquid acts in all directions.* Water will rush into a boat through a leak whether the leak is in the bottom or in the sides.

To show that the thrust on a surface increases with depth and at a given depth it is the same in all directions, we may use the arrangement shown in fig. 27. A thin rubber membrane is stretched across the mouth of a thistle tube *A*. *A* is connected by a rubber tube to a glass tube *B* of narrow bore. *D* is a pellet of mercury serving as index. If the membrane is pressed, *D* moves. The arrangement acts as a pressure gauge.

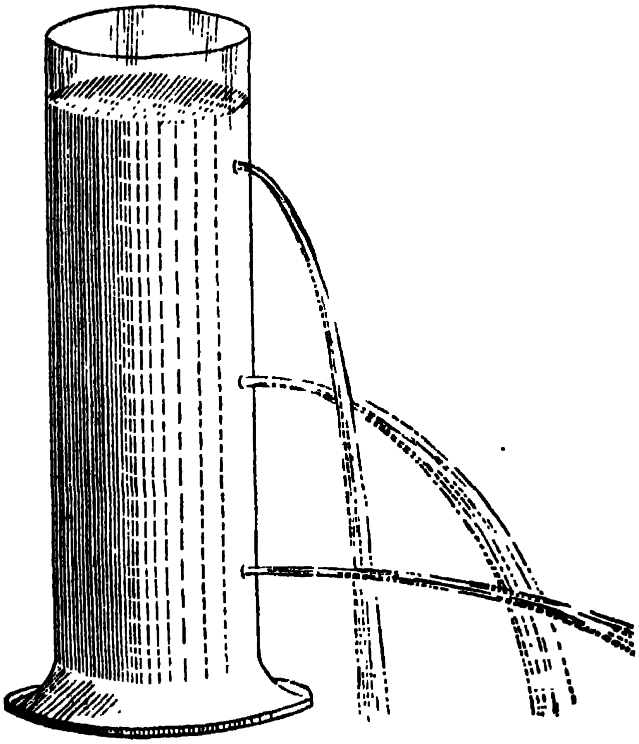


Fig. 26

Push *A*, mouth downwards, into a liquid. As the depth of immersion increases *D* moves more and more away from *A* showing that the force on the membrane increases with depth. Hold *A* at a fixed depth below the liquid, and so turn it that the centre of the membrane remains at the same depth. It will be found that *D* does not move. This shows that the force on the membrane is the same whatever the direction it faces, provided the depth of its centre below the surface remains unchanged.

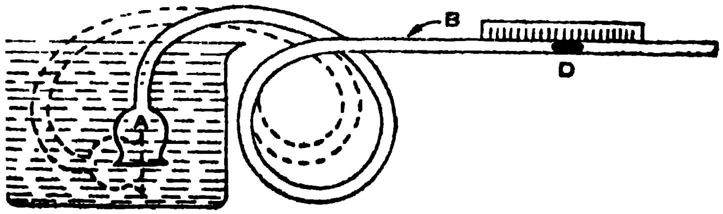


Fig. 27

(iv) **The magnitude of the thrust on any surface is equal to the weight of the liquid above the surface.** This can be demonstrated in the following way:

Take a cylinder open at both ends. Its lower end can be closed water-tight by a light glass or metal plate. When the cylinder, with the plate closing one end of it, is pushed into a liquid the upward force on the plate holds the plate in position (Fig. 28).

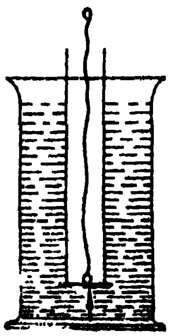


Fig. 28

Carefully pour liquid into the cylinder. The weight of the liquid will press down on the plate and gradually neutralise the upward force. The plate drops when the liquid in the cylinder reaches the same level as that outside. This clearly shows that the upward force exerted by a liquid upon a surface is equal to the weight of the liquid above the surface.

37. Pressure at a Depth in a Liquid

From the foregoing we conclude that

(i) *a liquid exerts a normal force (i.e. a thrust) on any surface with which it is in contact,*

(ii) *such forces act in all directions,*

(iii) *the magnitude of the force is equal to the weight of the liquid above the surface.*

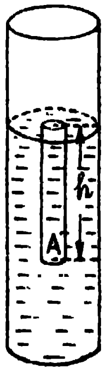


Fig. 29

Instead of referring to the total force acting on a surface, it is more convenient and often more profitable to refer to the pressure, i.e., the force per unit area.

To obtain an expression for the pressure at a point in a liquid, at a depth h below the free surface of the liquid, imagine a small horizontal area A surrounding the point. Consider the liquid contained in the vertical cylinder which has A as its base (Fig. 29). The vertical forces on the cylinder are (i) the upthrust on the base, and (ii) the weight of the liquid in the cylinder. These two forces are equal.

The thrust on the base = the weight of the cylinder
=volume of cylinder × density of liquid
$=h\ A \times d$, where d is the density.

The pressure P at the point is, therefore, given by

$$P = \frac{\text{thrust on the base}}{\text{area of the base}} = \frac{h\ A\ d}{A} = h\ d$$

or **Pressure = Depth × Density.**

If the depth is in centimetres and density in grams per c.c., pressure will be in 'gm wt. per sq. cm.' If the depth is in feet, and density in pounds per cubic foot, pressure will be in 'lb. wt. per sq. ft.' If necessary the units can be changed easily, e.g., from 'pounds per sq. foot' to 'pounds per sq. inch' etc.

An indirect way of stating pressure. Sometimes the pressure is expressed in an indirect way by mentioning only the liquid and the depth. The expression 'a pressure of 30 inches of mercury' means the hydrostatic pressure due to a column of mercury 30 inches high, i.e., the pressure at a depth of 30 inches in mercury.

Mercury has a specific gravity of 13·6. Hence its density is 13·6×62·4 lb. per cft. or (13·6×62·4)÷(12×12×12) lb. per cubic inch.

∴ Pressure of 30 inches of mercury

$$=\frac{30\text{ in.}\times 13\cdot6\times62\cdot4\text{ lb.}}{12\times12\times12\text{ cu. in.}}=14\cdot7\text{ lb. per sq. in. (nearly).}$$

A pressure of 100 ft. of water will be 100×62·4 lb. per sq. ft. or $\frac{100\times62\cdot4}{12\times12}$ lb. per sq. in.

38. Thrust on a Surface immersed in a Liquid

The magnitude of the thrust will be given by the product of the pressure and the area provided the pressure is constant over the area ; i.e., if the surface is horizontal. Otherwise,

Thrust=Area× *Average pressure* over the area.

The average pressure over a rectangular wall confining a liquid is the pressure at the mean depth. If a dam is a quarter of a mile long and water is 100 ft. deep, the thrust on the dam=area of the dam under water×pressure at 50 ft.

$$=(1320\times100)\text{ sq. ft.}\times50\text{ ft.}\times62\cdot4\text{ lb./cu. ft.}$$
$$=1\cdot84\times10^5\text{ tons nearly.}$$

Larger pressures at greater depths would require that dams be constructed to have a greater thickness at the base than at the top.

39. Some Results of Dependence of Liquid Pressure on Depth only

It should be clearly noted that pressure in a given liquid depends on the depth of the point in question below the free surface of the liquid. In engineering practice this depth is called the 'head' of the liquid. Pressure at a point does not depend on how far below the point the liquid extends. Nor does the pressure depend on the shape of the vessel containing the liquid, or the area of its cross-section.

Some consequences of this property of liquid pressure are the following :

(i) ***The pressure at all points in the same horizontal plane inside a liquid is the same.*** We shall have to use this result in many cases.

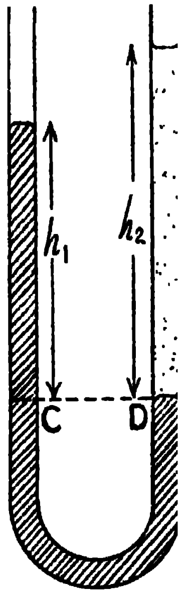


Fig. 30

When two liquids which do not mix are poured into the vertical limbs of a U-tube, the liquids stand at different heights in the two limbs. In Fig. 30, let h_1 and h_2 be respectively the heights above the common surface of separation, and d_1 and d_2, the densities of the corresponding liquids. Consider a horizontal plane CD drawn through the common surface. The pressure all over this plane CD (within the liquid) is the same. At C, the pressure is $h_1 d_1$. At D, it is $h_2 d_2$.

$$\therefore\ h_1 d_1 = h_2 d_2.$$

This result provides a method for comparing the densities of two liquids. Experiments based on it are described in § 40. Note that the result is not affected by the diameter of the tube. The two limbs of the tube may have different diameters.

(ii) *A liquid stands at the same height in communicating vessels* (Fig. 31). Notice that in a teapot water stands at the same level in the spout as in the body of the pot. This property of a liquid is sometimes expressed by saying that *a liquid seeks its own level.* Advantage is taken of this property of water in distributing water in a town from overhead tanks.

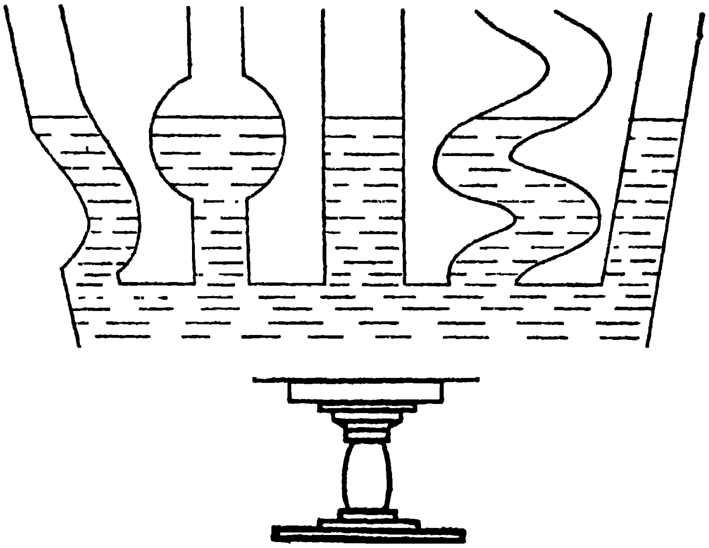


Fig. 31

Water-supply in a town. A town may be supplied with water from a distant source. The source of water may be a natural or artificial lake formed by putting a dam across a river. The channel, called an aqueduct, which carries the water from the lake to the town may pass under hills, fields, roads or streams. The water, after necessary purification, is forced into storage tanks, which are kept at a high elevation.

The distribution within the town is made through a system of pipes buried in the ground. Water is forced through these pipes by the pressure which results from the height at which the storage tanks are placed. The pressure at each faucet in the town depends upon the difference between the height of the storage tank and that of the faucet.

If a static condition prevailed water in buildings could reach the height of the storage tanks. But when water is flowing the friction in the pipes prevents the water from reaching the full height, i.e., from finding its own level.

40. Comparison of Densities of Liquids by U-tube and Hare's Apparatus

Experiment 22. *To compare the densities of oil and water by a U-tube.*

Apparatus. U-tube on stand, oil, water, mercury, scale.

Method. Mount the U-tube vertically on the stand. Fill its curved part with mercury (Fig. 32). Pour water into one limb and oil into the other *until the mercury levels in the two limbs are the same.* Measure the vertical heights of the columns of oil and water above the mercury in the two limbs. Let them be h_1 and h_2 respectively.

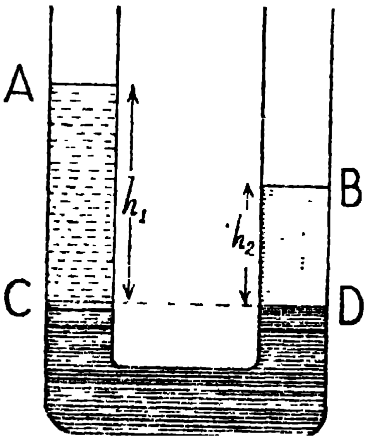


Fig. 32

Since mercury stands at the same level in both limbs, the pressure above it is the same. The pressure above the mercury in the limb containing oil$=h_1d_1$ where d_1 is the density of oil. The pressure in the other limb is $h_2 d_2$ where d_2 is the density of water.

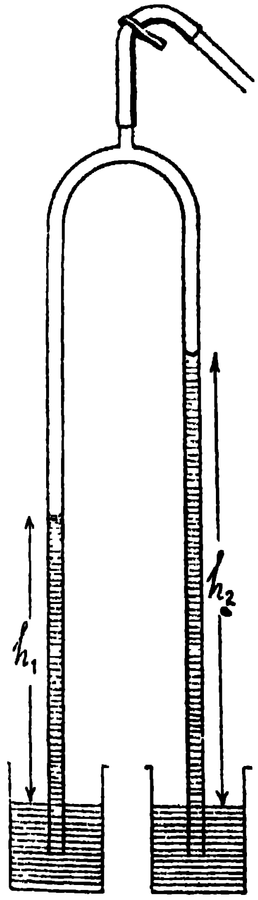


Fig. 33

$$\therefore\ h_1d_1=h_2d_2 \text{ or } d_1=d_2h_2/h_1$$

***Results*:**

Height of oil column, $h_1=$ cm.
Height of water column, $h_2=$ cm.
Density of water, $d_2=1$ gm per c.c.
$\therefore$ Density of oil$=h_2/h_1=\dots$ gm per c.c.

Experiment 23. *To compare the densities of two liquids by Hare's apparatus.*

Hare's apparatus (Fig. 33) consists of an inverted U-tube with a side tube carrying a piece of rubber tube and a pinch cock. The open ends of the limbs dip into two beakers containing the experimental liquids.

Apparatus. Hare's apparatus on vertical stand, beakers filled with experimental liquids, scale.

Method. Suck out air partially through the side tube of Hare's apparatus while its open ends dip in beakers containing the two liquids. The liquids rise to different heights in the limbs. Measure the height of each liquid column *from the free surface of the liquid in the beaker.* (The levels in the two beakers need not be the same.) Let the heights and densities be h_1, h_2 and d_1, d_2.

The pressure on the free surface of the liquid in a beaker is the atmospheric pressure P (See § 45). At the same level inside the tube the pressure is also P. But this is now made up of the pressure $h_1 d_1$ due to the liquid column and the pressure P' of the air in the apparatus.

$$\therefore \quad P = h_1 d_1 + P'$$

In the other limb, we have, similarly,

$$P = h_2 d_2 + P'$$

$$\therefore \quad h_1 d_1 = h_2 d_2.$$

(Results may be recorded as in the previous experiment.)

41. Pascal's Law of Transmission of Pressure by Liquids

From the fact that pressure within a free liquid depends only upon the depth and density of the liquid, Pascal, a French mathematician (1623-1662), drew a very important conclusion. It is known as Pascal's law and may be stated as follows :

Pressure applied anywhere to a body of confined liquid is transmitted by the liquid and acts undiminished at every point of the liquid and on the walls of the container.

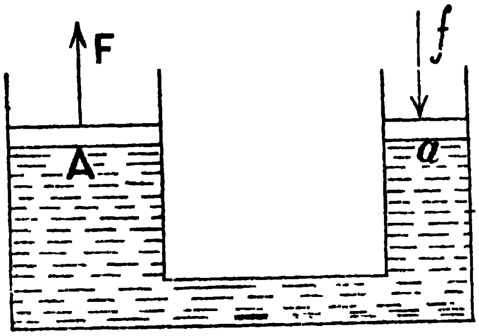


Fig. 34

With the help of Pascal's law *a small force can be transformed into a large one.* Consider the arrangement illustrated in Fig. 34. It is a confined body of liquid connecting two cylinders of areas a and A respectively, each fitted with a piston. If a force f is applied to the small piston, the increase of pressure on the liquid will be f/a. This pressure, transmitted undiminished by the liquid, will act on the larger piston and give rise to a force $F = A \times f/a$. The force f is thus multiplied by a factor A/a. If $a = 1$ sq. cm. and $A = 1000$ sq. cm. a force of 1 kg. applied to the small piston will cause a force of 1000 kg. to be applied to the large piston.

42. Hydraulic Press

The above principle is employed in the hydraulic press, a machine now in common use for a host of purposes, such as subjecting bales of cotton, jute etc. to enormous compressing forces, punching holes through metal plates, pressing metal sheets into shape, testing strength of iron beams, extracting oil from seeds etc.

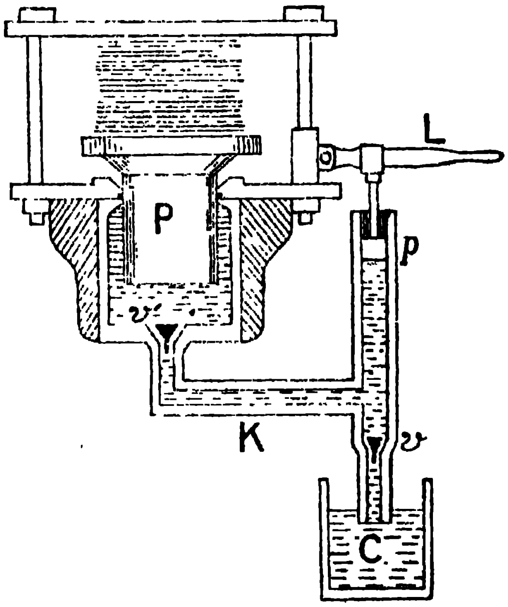


Fig. 35

Fig. 35 represents diagrammatically a hydraulic press in section. As the small piston *p* is raised, water from the cistern *C* enters the piston chamber through the valve *v*. As soon as the downstroke begins the valve *v* closes and the valve *v'* opens. The pressure applied on *p* is transmitted through the tube *K* to the large reservoir. There it acts on the large cylinder *P*. If the areas of cross section of *p* and *P* are *a* and *A* respectively, and the force on *p* is *f*, then the force *F* acting on *P* is $f \times A/a$.

The lever *L* also helps in multiplying the force. For if f' is the force applied to the lever, and *m* is the ratio of its longer arm to the shorter arm, then $f=mf'$. So, by exerting a force f' we get a force $F=mf' \times A/a$.

Hydraulic brakes used in automobiles, the hydraulic jack which raises heavy loads or the garage-lift which lifts automobiles provide further examples of application of Pascal's law.

43. Floating Bodies

According to Archimedes' principle a body, even if *partly* immersed in water, will experience an upward buoyant force equal to the weight of the liquid displaced. The downward force on the body is its weight. When the two forces are equal, the body will float. We may say that *a body will float when it displaces its own weight of the liquid.* When a floating body is placed in a liquid it will sink until it displaces its own weight of the liquid. If pushed further into the liquid, it will rise when released.

(i) Let *V* c.c. be the volume of a floating body of density *d* gm per c.c. *Let n be the fraction of its volume immersed in water when* *the body floats*. Then its weight is $V \times d$ gm wt. and the weight of of the displaced water is nV gm wt.

$$\therefore \quad Vd = nV$$

$$\text{or } d = n \text{ gm per c.c.}$$

If the liquid in which it floats has a density d', we shall have

$$Vd = nVd'$$

$$\text{or } d = nd'$$

The result may be expressed in words as follows :

If a solid floats in a liquid with a fraction n of its volume immersed, then the density of the solid is n times that of the liquid.

(ii) Let l_1 and l_2 be the distances to which a cylinder of cross-section A sinks in two liquids while floating in them. Let d_1 and d_2 be the respective densities of the liquids. By Archimedes' principle the weight w of the cylinder

$$= Al_1d_1 = Al_2d_2$$

$$\text{or } \frac{d_2}{d_1} = \frac{l_1}{l_2}$$

Now, if d_1 and l_1 refer to values for water, d_2/d_1 is the specific gravity of the other liquid. Hence we may say that

The specific gravity of a liquid is the ratio of the depth l_1, which the cylinder sinks in water, to the depth l_2, which it sinks in the liquid.

Problem. A wooden cylinder of uniform cross-section is 10 cm. long. It floats in water with 2 cm. above the surface. In a salt solution it floats with 3 cm. above the liquid surface. Find the density of the salt solution.

Let A sq. cm. be the area of cross-section of the cylinder, and d, the density of the solution in gm per c.c. Then the weight of the cylinder = weight of displaced water = weight of displaced liquid.

Now, weight of displaced water $= 8A$ gm wt.

weight of displaced liquid $= 7Ad$ gm wt.

$$\therefore \quad 7Ad = 8A \text{ or } d = 8/7.$$

(iii) **Sinking bodies.** It is easy to find the condition when a body will sink in a liquid. Let V be the volume of the body, supposed fully immersed in a liquid of density d'. Let d be the density of the body.

Then the weight of the body acting downwards $= Vd$.

The upthrust on it = weight of the displaced water $= Vd'$.

The body will sink if

its weight > the upthrust on it.

$$\text{i.e. } Vd > Vd'$$

$$\text{or } d > d'.$$

It is, therefore, clear that *a body will sink in a liquid if its density is greater than that of the liquid. If* $d < d'$ *the body floats.*

The density of iron is much greater than that of water. How does then an iron ship float in water ? This is due to the shape given to the iron in constructing the ship. The shape is such that the ship can displace a large volume of water, much more than what her iron would displace if immersed fully in water.

The average density of the human body is very close to that of water. When we breathe out air our body becomes denser than water and will sink. But when air is inhaled the body is lighter than water.

The water of the Dead Sea contains much salt in solution and is so dense that one does not sink in it.

(iv) **Hydrometers.** The varying depths to which a floating body sinks in liquids of different densities is utilized in the construction of commercial hydrometers. (See § 28). The following experiment may be done with profit to understand the action of hydrometers.

Experiment 24. *To construct a hydrometer from drinking straw.*

Apparatus : Drinking straw, molten wax, sand or lead shots, strip of graph paper.

Method : Select a piece of drinking straw with as uniform a section as you can get. Close one end with molten wax. Put some sand or a few lead shots into it so that the loaded straw can float in water with about three-fourths of its length immersed in water. Cut a narrow strip of graph paper and fix it along the tube. It will serve as a scale.

Float the tube in water and note where the water stands on your paper scale. Mark this place as 1000. Then float it in a solution of copper sulphate or brine of known specific gravity. Mark the place on the scale where the liquid stands, and put against it the value of the specific gravity of the liquid multiplied by 1000.

Count the number of divisions on the paper scale lying between these two marks. Let it be n. Let N be the difference between the readings at the two marks. Then each division on the paper scale represents a change of specific gravity of N/n units on the scale that the specific gravity of water is 1000. The values increase downwards.

Now measure the specific gravity of some other solution with the hydrometer you have so constructed.

(v) **Floating of Ships.** The modern ship may be regarded as a huge hydrometer. Though made of a material heavier than water, she does not sink because of the air she contains—much like a hydrometer made of glass. A ship has special lines painted on the outside, which mark the level beyond which she must not sink when loaded. The lines are marked with letters, such as FW, IS, S, W etc. FW stands for 'fresh water', IS for 'Indian Ocean in Summer', S and W for 'Summer' and 'Winter' in other seas. The W line is lowest and the FW line is highest because a ship of given weight sinks more in fresh water (which is lighter) than in the cold saline water of a winter sea (which is denser).

The lines are in a way analogous to the graduations on a hydrometer. They are called Plimsoll lines after Samuel Plimsoll who, in 1876, got an Act of Parliament passed restricting loading of ships beyond these lines.

When we say that a ship has a displacement of 10,000 tons we mean that the weight of the ship and cargo together will be 10,000 tons to sink her to the Plimsoll line.

(vi) **Balloons.** When a piece of cork immersed in water is released, it floats up due to the upthrust exerted on it by the water. Similarly, when a body lighter than air is released in air, it will float up. Archimedes' principle applies both to liquids and gases.

The balloon acts on this principle. Its bag or envelope is made of two or three layers of thin cotton or silk impregnated with rubber so as to make it light, strong and gas-tight. Passengers, equipment and ballast, if any, are placed in a gondola suspended from the envelope by light cords. The gas may be hydrogen or helium. Helium is non-inflammable and is preferred to hydrogen, though the latter has a greater lifting power as it is lighter.

A balloon is not completely filled at the start. As it ascends the outside pressure diminishes, and the gas in the bag expands. The balloon rises to a height where its total weight equals the weight of the air it displaces. To come down, some gas is allowed to escape through a valve.

Exercises

1. Distinguish between pressure and thrust. Express a pressure of 1 ton per sq. ft. in pounds per sq. in. (*Ans* : 15·5).

What do you understand by the expression 'pressure at a point'?

2. Describe simple experiments to demonstrate any two of the following :

(a) A liquid exerts a normal thrust on any surface with which it is in contact.

(b) The thrust exerted by a liquid increases with depth.

(c) The thrust due to a liquid acts in all directions.

3. How would you establish the relation that in a liquid, pressure = depth × density ?

4. What do you understand by the expression 'a pressure of 30 inches of mercury' ? Express it in pounds per sq. in., given that the sp. gr. of mercury = 13·6 and 1 cft. of water weighs 62·4 lb. (*Ans*: 14·7)

5. The water on the upstream side of a dam is 120 ft. deep. If the river bed is of rectangular section and the dam is half-a-mile long, find the thrust on the dam in tons. (*Ans*: $5{\cdot}3 \times 10^5$).

Why is a dam thicker at the base ?

6. How can you compare the densities of two liquids with the help of a U-tube ?

Is it necessary that the limbs of the tube should have the same area of cross section ? Explain your answer.

7. What advantage does Hare's apparatus offer over the U-tube for the comparison of densities of liquids ?

Why, in Hare's apparatus, is it necessary to measure the height of a liquid column from the free surface of the liquid in the beaker ?

8. State Pascal's law, and illustrate how a small force can be transformed into a large one with its help.

9. Draw a sketch to illustrate the action of a hydraulic press, and explain how it acts.

Mention some of its uses.

10. What are the conditions that a body will (i) float, (ii) sink, in a liquid ?

11. Calculate the pressure at a depth of 76 cm. of mercury. What height of water will exert the same pressure ? (Sp. gr. of mercury = 13·6). (*Ans*: 1034 gm. wt. per sq. cm; 1034 cm.)

12. The level of water in the storage tank of a water-supply system is 80 ft. above the ground. If the loss of pressure head due to friction etc. is 50 ft., what is the pressure at a faucet 20 ft. above the ground ? (*Ans* : 10 ft. of water, or 624 lb. per sq. ft).

By lowering the level of a tap you can get more water out of it. Why ?

13. The bend of a U-tube is filled with mercury. Enough brine of sp. gr. 1·10 is poured into one limb to create a difference of 1 cm in the levels of mercury in the two limbs. What is the height of the column of brine ? (Sp. gr. of mercury =13·6).
(*Ans* : 12·4 cm.)

15. If the sp. gr. of ice is 0·917 what fraction of its volume will jut out when ice floats on water ? (*Ans* : 0·083)

16. A uniform stick sinks 0·8 of its length in a liquid and 0·9 in water. Find the density of the liquid. (*Ans* : 9/8 gm per c.c.)

17. A straight stick is 1 sq. in. in cross section. It is pushed 4 ft. into water. What is the upthrust on the stick ? If the stick weighs one pound, how much force will be needed to hold the stick in position ? What is the direction of this force ?
(*Ans* : 1·73 lb. wt. and 0·73 lb. wt. nearly; downwards.)

18. A solid floats with 1/6 of its volume above water. What fraction of its volume will project if it floats in a liquid of density 1·2 gm per c.c. ? (*Ans* : 11/36)

19. The specific gravity of ice is 0·916, and that of sea-water is 1·025. If the volume of an ice-berg projecting out of water is 1000 cubic feet, how much of it is submerged ? (*Ans* : 8404 cft)

20. A mine of volume 200 litres and mean specific gravity 0·95 is held under water by a light chain fastened to the sea-bed. Calculate the upthrust on the mine when the specific gravity of sea-water is 1·02. What is the tension in the chain ?
(*Ans* : 204 kgm. wt.; 14 kgm wt.)

CHAPTER 5

AIR PRESSURE AND ITS MEASUREMENT

44. Atmosphere

Everything on the earth's surface is submerged in an ocean of air. This gaseous envelope surrounding the earth is called the *atmosphere.* It is not possible to say precisely how far above the earth the atmosphere extends. The air is densest at sea-level, gets thinner and thinner the higher up we go and finally thins out into interstellar space. Experiments with radio waves show that small amounts of air exist even at a height of more than two hundred miles. Much information in this regard has been and is being collected through rockets shot high into the air and through the artificial satellites. But very little of this information has so far been released to the public.

The molecules of air move about at terrific speeds. The gravitational attraction of the earth prevents them from flying away into interstellar space and being lost to the earth. The force of gravity is directed towards the centre of the earth and is stronger at lower heights above the surface than higher up. Hence air molecules tend to crowd together near the surface of the earth. Complete crowding is prevented by the motion of the molecules. The result is that there are more molecules crowded in a given volume near the surface of the earth than above it. This is why the atmosphere gets thinner and thinner as we go higher up.

The total weight of the atmosphere has been estimated to be 5×10^{15} tons. About 50% of it lies within 3½ miles of the surface and 99% within 20 miles.

45. Atmospheric Pressure

The atmosphere exerts a pressure on every thing which is in contact with the air. This pressure, called the *atmospheric pressure,* is equal to the weight of a column of air contained in an imaginary cylinder of unit cross-section and extending up to the top of the atmosphere. At the sea-level the magnitude of this pressure has been found to be equal to a weight of 14·7 lb. per square inch.

46. Demonstration of Atmospheric Pressure

The existence of atmospheric pressure can be demonstrated in various ways. Like liquid pressure it acts in all directions and is perpendicular to any surface with which the air is in contact. If a surface has air on both sides of it there is no unbalanced force due to atmospheric pressure acting on the surface. The thrusts exerted by the atmosphere on the two sides of the surface are equal and opposite. To demonstrate the effect of atmospheric pressure on a surface, *the air must be fully or partially removed from one side.*

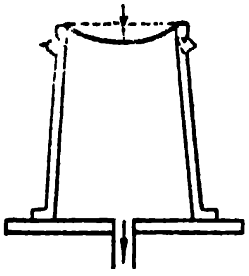


Fig. 36

(i) A thick walled glass vessel (Fig. 36) open at both ends is closed at one end by a stretched rubber membrane. The other end, which is well ground, is placed on the receiver of an air pump. As the air inside the vessel is gradually pumped out, the membrane sags more and more. It may eventually burst.

Before the air was pumped out the membrane had been subjected to the same pressure due to air on both sides. As air was removed from one side, it gradually yielded under the pressure exerted on the other side.

If the receiver were inclined at various angles while the pump was running, there would be no change in the sagging of the membrane. This would show that the pressure is exerted in all directions.

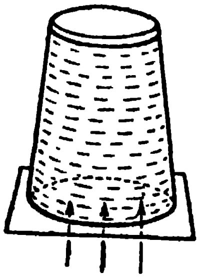


Fig. 37

(ii) A glass tumbler, full to the brim with water, is covered with a thin metal disc or a piece of card-board, leaving no air inside. The tumbler is then inverted while the card is still pressed to it (Fig. 37). On removing the support from the card neither the card nor the water will be found to drop. They are kept in position by the upward pressure of the air acting on the card. For the experiment to succeed the card should touch the rim at all points.

We shall see later that air can, in this way, support a column of water about 34 feet high, or a column of mercury 76 cm in height, and that the height of such a column provides a measure of the atmospheric pressure.

(iii) **Magdeburg Hemispheres.** A very impressive experiment to demonstrate the effect of atmospheric pressure was performed by Otto von Guericke, a Prussian scientist, before Emperor Ferdinand in 1654. Two copper hemispheres, about 22 inches in diameter, were placed together to form a sphere as shown in Fig. 38. A ring of leather soaked in oil and wax was set between them to make an air-tight joint. When the air was pumped out of the sphere two teams of eight horses each were required to pull the hemispheres apart. It is easy to calculate that the force holding the hemispheres together amounted to about three tons.

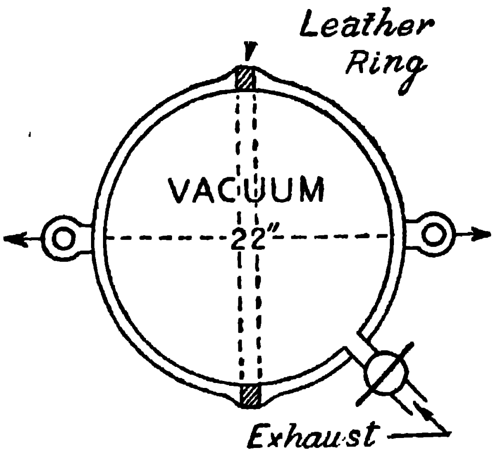


Fig. 38

Class experiments are generally shown with hemispheres as in Fig. 39. The surfaces of the hemispheres which come into mutual contact are well ground. A thin layer of vacuum grease makes the joint air-tight. For a sphere four inches in diameter, the total force pressing the two together is about 180 lb. when they are evacuated.

(iv) **Collapsing can experiment.** The existence of air pressure acting in all directions can also be shown as follows. Pour a small quantity of water in a gallon can and boil the water vigorously for a few minutes until the steam replaces most of the air in the vessel. If the mouth of the can is then tightly closed with a rubber stopper and the can allowed to cool, the steam inside will condense and pressure inside the can will fall to a low value. Unless the can is very stout, the greater pressure outside will be enough to crush it.

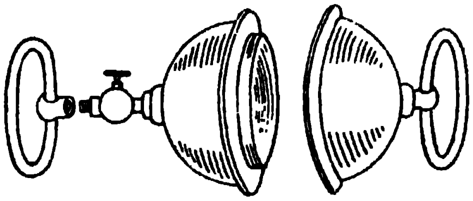


Fig. 39

Further examples of the action of atmospheric pressure are provided by some very well known phenomena. In breathing, our chest muscles pull the ribs upward and outward while the diaphragm flattens. This increases the size of the chest cavity and reduces the pressure around the lungs. The atmospheric pressure then forces air into the lungs. In drinking water or sipping a cold drink with a straw, we diminish the air pressure inside the tube and in our mouth cavity by increasing the volume of the latter. The liquid is then forced into our mouth by the higher pressure outside (Fig. 41).

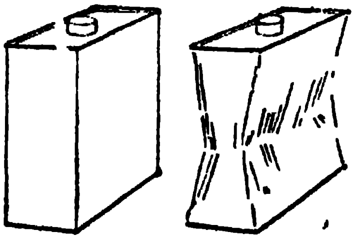


Fig. 40

When the lever of a self-filling fountain pen (Fig. 42) is raised, it presses a rubber bag within the pen and forces some air out of the bag. When the lever is lowered, the bag regains its old volume ; but the pressure in it is less than before. Atmospheric pressure then forces the ink into the bag.

47. Height of Mercury Column Atmospheric Pressure can support

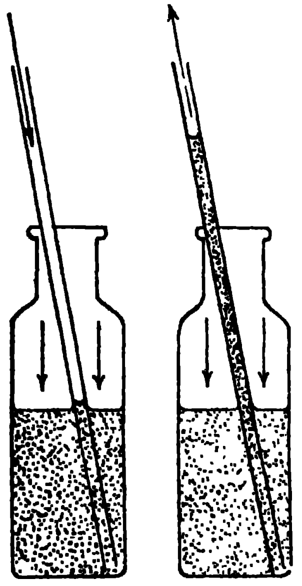


Fig. 41

From the examples cited above it would appear that Nature abhors vacuum, and, with the material readily available, she immediately fills any space in which a partial vacuum has been created. There is, however, a limit to her ability of doing so as would be clear from the following experiment.

Take a glass tube about a metre long and dip one end of it vertically in a large bowl of mercury (Fig. 43). Connect the other end of the tube to an air pump through a piece of pressure tubing (i.e., thick walled rubber tube). As the pump sucks air out of the tube, atmospheric pressure, which is pressing on the surface of the mercury in the bowl, forces mercury into the tube. If it is a good pump and you are not doing the experiment in a hill station you will find that the mercury rises to a height of about 30 inches and no more. Water under the same condition will rise to a height of about 30 feet. The column of liquid is supported by atmospheric pressure.

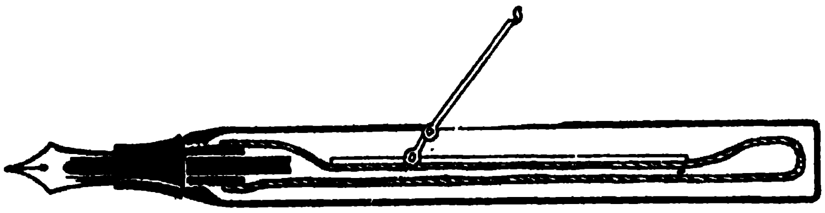


Fig. 42

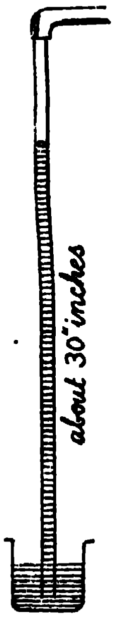


Fig. 43

Torricelli's Experiment. Torricelli (1662), an Italian scientist and a pupil of Galileo, was the first to conclude that it was atmospheric pressure which supported a column of mercury in an evacuated tube. He completely filled a tube about a metre long with mercury and inverted it in a vessel of mercury

(Fig. 44). The mercury in the tube came down to a height of about 76 cm above the mercury in the vessel. It fell no further because atmospheric pressure, pressing on the free surface of mercury in the wider vessel, was able to support the weight of the column of mercury. The condition for the balance is

Pressure exerted by the atmosphere =Hydrostatic pressure exerted by the column of liquid.

In order to test Torricelli's theory that the mercury in the tube was supported by atmospheric pressure, Pascal had the experiment performed at a height of about 5000 ft. above sea-level. If atmospheric pressure at a place is due to the weight of the air above the place, the pressure should decrease as we go higher. The height of mercury supported in the tube should therefore be less the higher we go. Pascal found this to be true.

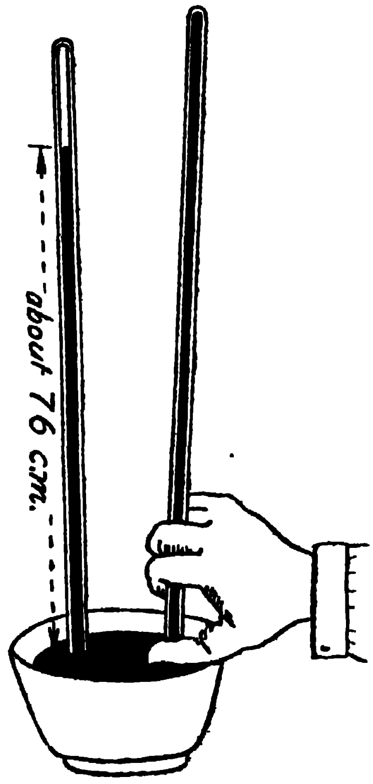


Fig. 44

The space above the mercury column (in Torricelli's experiment) is a vacuum since air has been excluded from the tube. It is known as *Torricellian vacuum*. If the tube is gradually inclined, more mercury goes into it, but mercury in the tube remains at the same vertical height above the mercury outside (Fig. 45). If the tube is tilted enough to make the mercury strike the end of the tube, it does so with a sharp click, showing that there is no air inside.

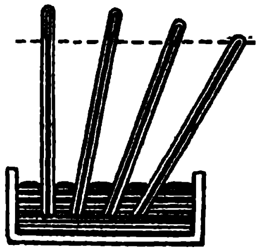


Fig. 45

To prove conclusively that the mercury in the tube is supported by the pressure of air acting on the surface of mercury in the wider vessel, we can perform the following experiment. Completely fill a glass tube, about a metre long, with mercury and invert it over some mercury in a glass jar (Fig. 46). The jar can be closed air-tight with a rubber stopper. A bent glass tube makes it possible to connect the air inside the jar with a suction air-pump. As the pump works the pressure of air inside the jar diminishes and the mercury column falls. On readmitting air mercury again rises in the tube.

The height to which the mercury rises is independent of the diameter of the tube provided it is not very narrow. In very narrow tubes another effect, due to what is called surface tension, slightly reduces the height.

48. Measure of Atmospheric Pressure

Atmospheric pressure is measured by the height of the mercury column it can support. It varies with the height of a place above the sea-level. Even at a given place it varies slightly according to the prevailing weather condition. For convenience of reference a standard value for atmospheric pressure has been defined as follows :

A pressure of one standard atmosphere is the hydrostatic pressure exerted by a column of mercury 76 *cm high at* 0°*C*, 45° *latitude and mean sea-level.** (Also see the section marked 'An indirect way of stating pressure' under § 37.)

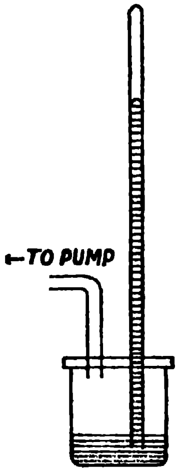


Fig. 46

A pressure of one atmosphere may be expressed in other units as follows :

1 atmospheric pressure = hydrostatic pressure exerted by 76 cm. of mercury.

$$=76\times13{\cdot}6=1034 \text{ gm wt. per sq. cm. } (\S 37)$$

$$=1034\times980=1{\cdot}013\times10^6 \text{ dynes per sq. cm } (\S 66)$$

If we take atmospheric pressure to be equal to 30 inches of mercury, which is the value we accept while using the British system of units, we shall have

Pressure of 1 atmosphere = hydrostatic pressure due to 30 inches of mercury

$$= 30 \text{ in.} \times \frac{13{\cdot}6\times62{\cdot}4 \text{ lb. wt.}}{12\times12\times12 \text{ in.}^3}$$

$$=14{\cdot}7 \text{ lb. wt. per sq. inch.}$$

Meteorologists usually express atmospheric pressure in *millibars*. As one millibar is equal to a pressure of 1000 dynes per sq.cm., a pressure of one atmosphere is equal to 1013 millibars (i.e., 1·013 bars).

The hydrostatic pressure due to 30 inches of mercury is equal to the pressure of 30×13·6 inches, i.e., about 34 feet of water. Atmospheric pressure can therefore support a column of water 34 ft. high. This is known as the height of the *water-barometer*.

---

* Mean sea-level is defined by an Act of the British Parliament as the half-way level between the average high and low tides at Newlyn in Cornwall.

49. The Barometer

A barometer is a device for measuring atmospheric pressure. It has already been seen that atmospheric pressure can support a column of liquid. So the measurement of atmospheric pressure consists in measuring the height of a liquid column supported by it. The liquid chosen is mercury.

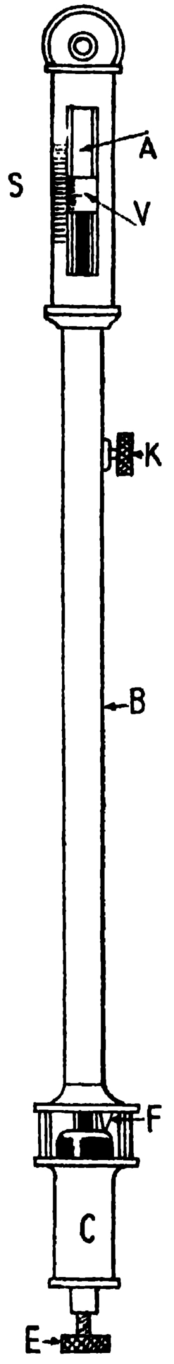


Fig. 47

A. Fortin's barometer. The most reliable barometer is the one devised by Fortin. It acts on the principle of Torricelli's experiment. In Fig. 47, *C* is a mercury cistern, in communication with the atmosphere, over which a glass tube *A* is inverted. The two are enclosed in a metal tube *B* on which a scale *S* is etched. There is also a vernier *V* which can slide in a slot made in the metal tube. The tube is so slotted as to make the highest part of the mercury column visible. The zero of the scale starts from the tip of an ivory pointer *F* which projects downwards from the ceiling of the cistern. The lower part of the instrument is shown magnified in Fig. 48. The surface of the mercury in the cistern can be adjusted by the screw *E* so that the tip of the ivory pointer *F* just touches it. The bottom of the cistern is made of leather.

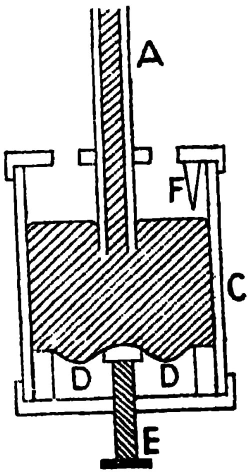


Fig. 48

Experiment 25. *To read a Fortin's barometer:*

(i) Find the vernier constant of the scale *S*.

(ii) Adjust the mercury surface by the screw *E* so that the ivory pointer *F* just touches it. This happens when the image of the tip of the pointer in the mercury coincides with the tip of the pointer itself.

(iii) Keep your eye at the level of the convex top of the mercury column. Slide the vernier by the screw *K* and so set it that its lower edge is tangential to the curved mercury surface.

(iv) Take the reading of the main scale and also of the vernier.

Note also the temperature from the thermometer attached to the instrument. Repeat several times by resetting the mercury level and the vernier. Record as follows :

Vernier constant = ...cm.

No. of readings	Main scale reading	Vernier reading	Total	Average
1	...	...	...	...
2	...	...	...	...
3	...	...	...	...

The atmospheric pressure = ...cms. of mercury at....°C

Similarly, take the reading of the barometer in inches.

B. The siphon barometer

It is very simple in construction and is portable, but it is not so reliable as Fortin's. It is (Fig. 49) a U-tube with one arm about 90 cm. long with the upper end closed. The other arm is short and open to the atmosphere. The tube is set up on a wooden board fitted with a scale. The difference of the mercury levels in the two arms gives the barometric height, which can be read from the scale. Unless very carefully constructed a siphon barometer may have traces of air and vapour in its Torricellian space. They lower the reading.

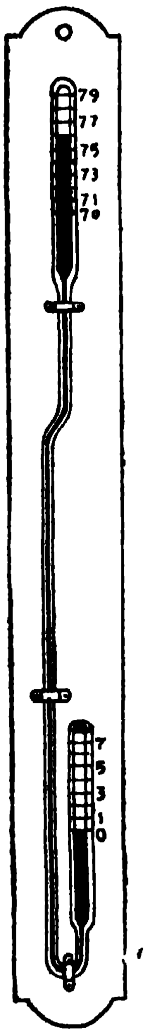


Fig. 49

C. Aneroid Barometer

This type does not require any liquid for its working ('aneroid' means without liquid) and has the great advantage of ruggedness and portability (Fig. 50). It consists of a partially evacuated metal box with a flexible corrugated top. A box of this type will collapse under pressure of air. This is prevented by a stiff steel spring *S* which pulls the top upwards.

The free end of the spring *S* is connected to a lever *L* which has its fulcrum at *F*. The other end of *L* is connected to a small chain *C* which is wrapped round a spindle *D* carrying a pointer *P* which moves over a dial. The movement of the corrugated top due to changes in the atmospheric pressure is magnified by the lever *L* and transmitted to the pointer *P*. The dial of the aneroid is calibrated by comparison with a standard mercury barometer.

Altimeters. Since the atmospheric pressure diminishes as we go up to higher altitudes, the barometer may conveniently be used to determine elevation. Altimeters are simply aneroid barometers with an altitude scale attached. The outer scale in fig. 50 is an altitude scale, each unit representing 1000 ft. Altimeters are very handy and sometimes small enough to be carried in a pocket. They can be made so sensitive as to indicate a change in elevation of 1 ft.

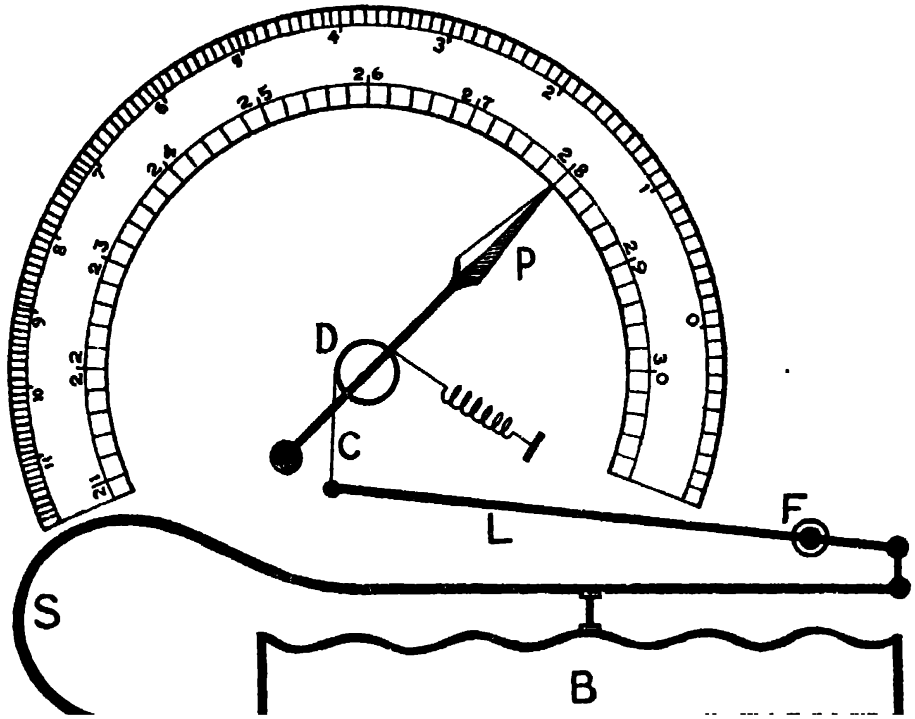


Fig. 50

At sea-level the atmospheric pressure changes approximately by 1 mm. of mercury for every 11-metre increase of height or 0·1 inch per 90 ft. of ascent.

50. Boyle's Law

Gases are highly compressible. Consider a mass of gas confined in a cylinder by an air-tight piston. If the pressure on the gas is increased by pushing the piston inwards, the volume of the gas decreases. As the pressure is reduced, the volume increases.

The relation between the pressure (*P*) and volume (*V*) of a fixed mass of gas was discovered by Robert Boyle (1627-1691), an English philosopher. It is known as **Boyle's Law** and may be stated as follows ;

For a fixed mass of gas at a given temperature, the volume *varies inversely* as the pressure, i.e., the product of the pressure and volume is constant.

In symbols, $PV = \textbf{constant} = K$.

If P_1 and V_1 are the initial values of the pressure and volume and P_2 and V_2, the final values we shall have

$$P_1 V_1 = P_2 V_2 = K$$

The value of the constant K depends upon (i) the mass of the gas, (ii) its temperature and (iii) the units in which P and V are expressed.

Density and pressure. Let D_1 and D_2 be the densities of a gas at pressures P_1 and P_2, and V_1 and V_2 the volumes of the same mass m of the gas at these pressures, the temperature remaining constant. Then

$$m = V_1 D_1 = V_2 D_2$$

By Boyle's law $P_1 V_1 = P_2 V_2$

Dividing the first relation by the second, we have

$$\frac{D_1}{P_1} = \frac{D_2}{P_2}$$

which means that *the density of a gas is proportional to its pressure, so long as the temperature is constant.*

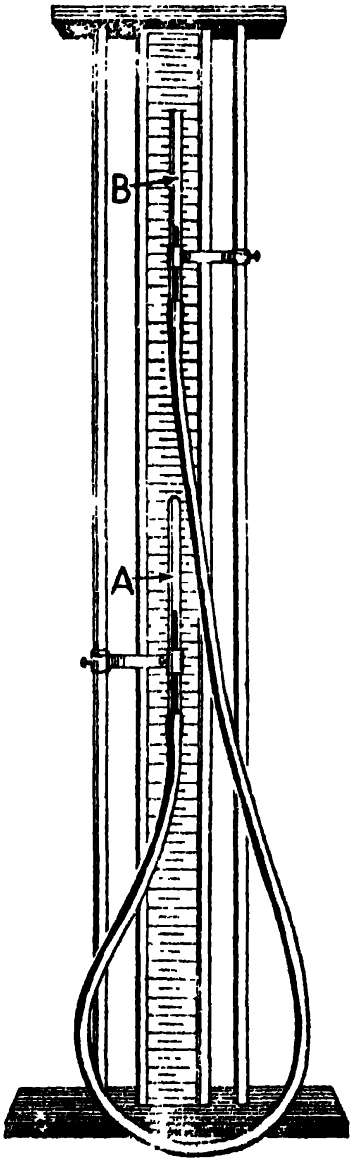


Fig. 51

51. Experimental Verification of Boyle's Law

Two glass tubes A and B are connected through a thickwalled rubber tube (Fig. 51). They are attached to a vertical frame and may be raised or lowered. A scale is fixed on the board between the tubes. The tube A contains the experimental gas and may have its upper end closed by a stop cock. A scale is also etched on the tube A by which the volume of the gas can be measured. The tube B is open to the atmosphere. A portion of B, the rubber tube and also a portion of A are filled with mercury.

To verify Boyle's law observations should be recorded for

pressures (i) equal to atmospheric, (ii) greater than atmospheric and (iii) less than atmospheric. If there is a stop cock above A, it should be perfectly closed so that the mass of the gas may not change during the experiment.

(i) The tubes A and B are adjusted so that the mercury stands at the same level in both. This means that the pressure of the gas in A is equal to the atmospheric pressure. Read this pressure from a barometer. Let it be h cm. of mercury. The volume of the gas in this condition is read off from the etched scale.

(ii) The tube B is now gradually raised. Mercury in B will stand at a higher level than that in A. For any position of B let the difference in mercury levels in the two tubes be x cm. The surface of the mercury in B is at atmospheric pressure, i.e., h cm. of mercury. Therefore, the pressure of the gas in A is equal to $(h+x)$ cm. of mercury. The volume of the gas is read from the graduations on A. The tube B is set at different heights and the observations repeated. It will be seen that for every position of B, the product of the pressure, $(h+x)$, and its corresponding volume is a constant.

(iii) The tube B is now lowered. The mercury in B will stand at a lower level than that in A. Let the difference of the mercury levels in the two tubes be l cm. It is clear that the pressure of the gas in A is now $(h-l)$ cm. of mercury. The volume is determined in the usual way. Several observations are taken for pressures less than atmospheric. Here also for all positions of B, the product of the pressure and the corresponding volume is a constant.

It will be seen that the product of the pressure and its corresponding volume has the same constant value in all the three cases.

It is assumed that the room temperature has not changed during the experiment.

Exercises

1. What is meant by 'atmosphere' and 'atmospheric pressure'? Describe two experiments to show that air exerts pressure.

2. How would you show that atmospheric pressure can support a column of mercury about 30 inches high near sea-level?

What proof would you offer to support that the effect is due to atmospheric pressure?

3. What is atmospheric pressure due to? What analogy has it with liquid pressure?

What do you understand by the statement that atmospheric pressure is 76 cm. of mercury ?

4. Express a pressure of one atmosphere in terms of the height of (i) a water column, (ii) a column of sea water of density 1·05 gm per c.c.

5. Draw a simple sketch showing the essential parts of a Fortin's barometer, and label the parts. Describe how the barometer works.

6. Examine the barometer in your laboratory and say what will happen to it if you take it to a height of 15,000 ft. where the air pressure is about 20 inches of mercury.

7. State Boyle's Law. How would you verify it for pressures (i) lower than atmospheric, (ii) higher than atmospheric ?

8. 190 c.c. of an enclosed mass of gas exert a pressure of 1000 mm. of mercury. If the gas is brought back to the normal atmospheric pressure (760 mm. of mercury), what will be its new volume? (*Ans* : 250 c.c.)

9. The volume of an enclosed gas in the closed limb of a Boyle's Law apparatus is 900 c.c. when the mercury column in the open limb stands 5 cm. higher than that in the closed limb. Find the volume of the enclosed gas when (i) the level in the open limb is higher than that in the closed limb by 14 cm., (ii) the level is lower by 16 cm. The atmospheric pressure is 76 cm. of mercury. (*Ans* : 810 c.c.; 1215 c.c.)

10. The volume of an air bubble is 1·5 c.c. at a depth of 230 ft. under sea-water (sp. gr. 1·03). Find the volume of the bubble when it rises to the surface. The height of the water barometer is 34 ft. at the place. (*Ans* : 12 c.c. nearly.)

11. An open bottle is immersed upside down in sea-water (sp. gr. 1·03). To what depth is it to be immersed so that 1/4 of its volume is occupied by water? The atmospheric pressure is 30 inches of mercury. (*Ans* : 11 ft.)

12. The density of air at 0°C and standard atmospheric pressure is ·00129 gm per c.c. What will be the density if the pressure falls to 74 cm. of mercury ?

If a globe has a capacity of 15 litres how much air will it lose at the lower pressure ? (*Ans* : ·00126 gm per c.c.; ·45 gm.)

CHAPTER 6

SIPHON, PUMPS AND MANOMETERS

52. The Siphon

A siphon is a simple device for transferring a liquid from one vessel to another at a lower level without tilting the vessel. When it is not possible or convenient to lift a vessel of liquid, such as a petrol tank, to empty it, we often use a siphon. We also use it to draw off the upper layers of a liquid without disturbing the lower layers.

In its simplest form a siphon (Fig. 52) consists of an inverted U-tube of unequal limbs, *completely filled with liquid.* The shorter limb dips into the liquid to be transferred ; the longer limb leads outside and extends below the level of the liquid to be drained.

The action of the siphon may be understood as follows. Since the siphon tube is full of liquid, the liquid must flow through the tube if the force pushing on the liquid at one end of the tube is greater than that at the other end. The upward pressure at A is equal to atmospheric pressure minus the downward pressure due to the liquid column DA. The upward pressure at B is equal to the atmospheric pressure minus the downward pressure due to the liquid column EB. Hence the pressure at A is greater than the pressure at B by an amount equal to the pressure due to the liquid column $FB=EB-DA$.

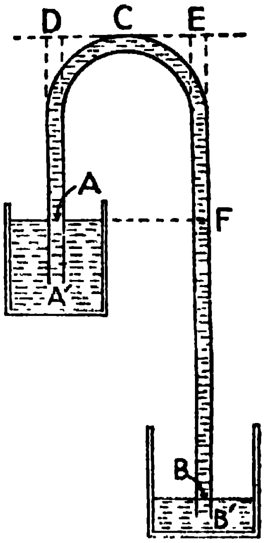


Fig. 52

The siphon will cease to act when the liquid is at the same level in the two vessels, for then $FB=0$ and the forces acting at the two ends of the siphon are equal and opposite. If the liquid is water it will also cease to act if the bend C is more than 34 ft. above the surface of water in the higher vessel, for then the atmospheric pressure will be unable to lift the water up to the bend C. A siphon will not work in vacuum, for it is the atmospheric pressure which pushes the liquid through.

53. Water Pumps

Water pumps are devices for raising water from a lower to a higher level. Since liquids possess a natural tendency to flow from a higher to a lower level, the lifting is possible only at the expense of energy supplied from outside.

There are different types of pumps. Each type, however, has valves which allow flow of liquid in one direction only. The difference between the types lies mainly in the location of these valves. It should be clearly understood that by suction alone water can never be lifted if the depth of water from the level of the pump is greater than the height of the water barometer (34 ft). In practice, however, due to leakage and to vapour pressure such pumps would not work if the height is greater than 30 ft. or so.

(i) The Common or the Suction Pump

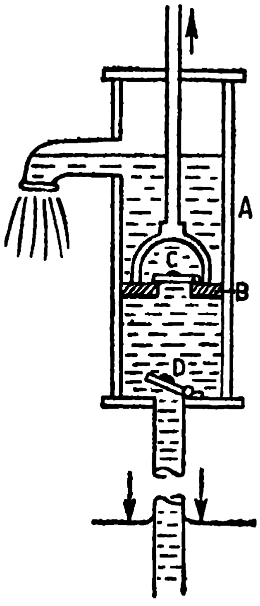


Fig. 53

The construction and the action of the pump will be clear from Fig. 53. A piston *B*, with a valve *C* at its head, fits airtight into the cylinder *A*. The valve *C* opens upward and so allows flow only in the upward direction through the piston. From the bottom of the cylinder a pipe goes down into the reservoir from which water is to be pumped up. At the junction of the pipe and the cylinder there is another valve *D* which also opens upwards.

Suppose the piston occupies the lowermost position at the start. During the upstroke, a partial vacuum is created in *A* between the piston head and the valve *D*. The higher pressure from below lifts the valve *D* and air from inside the pipe enters the cylinder. As a result a partial vacuum is created in the pipe also.

Atmospheric pressure now forces some water up the pipe. As the pump is gradually worked water rises more and more into the pipe and ultimately enters the cylinder. In this condition each downstroke will cause some water to pass through *C* and collect above the piston head. During the next upstroke this water will flow out through the spout, while more water will flow from the pipe into the cylinder. This intermittent outflow of water at each upstroke will continue so long as the pump is worked.

Priming. The valve *C* is generally not airtight, particularly when the pump has been in use for some time. As a result the degree of vacuum within the cylinder is not sufficient to make the valve *D* open. To avoid this difficulty a part of the cylinder above *C* is filled with water from outside to make the valve *C* air-tight. Suction is now effected easily. This operation is called 'priming'.

(ii) **The lift pump**

A suction pump, as described above, cannot raise water more than 30 ft. or so for reasons stated earlier. To lift water to greater heights a lift pump may be used. The mechanism will be clear from Fig. 54. nI place of the spout a delivery tube *F* is attached to the cylinder. At the junction there is a valve *E* which opens outward, i.e., in the direction of flow. The other valves *C* and *D* are in the same position as in a suction pump. The principle of action is as in the suction pump. When the cylinder is filled with water, every upstroke forces some water into *F* through *E*. The water gradually rises into *F* and is delivered at the desired height. Note that here also atmospheric pressure forces the water from the reservoir into the cylinder of the pump. Hence the depth of water in the reservoir from the level of the pump should not exceed about 30 ft. Extra energy is needed at each upstroke to push the water into the pipe *F* against the hydrostatic pressure of the liquid column in *F*.

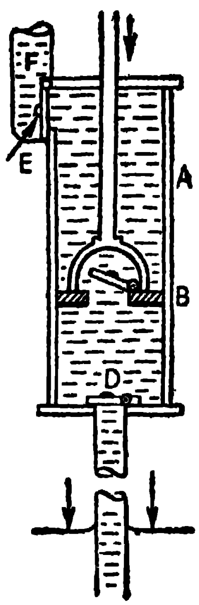


Fig. 54

(iii) **The force pump**

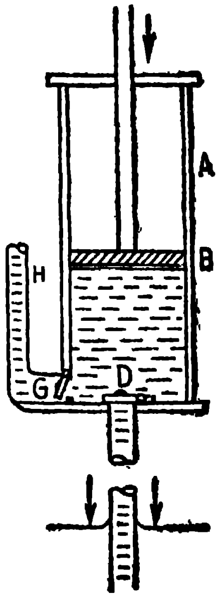


Fig. 55

The force pump (Fig. 55) can also raise water to heights greater than 30 ft. But here the piston is a solid rod and the delivery pipe *H* is connected at the bottom of the cylinder through a valve *G* which opens only in the direction of flow. The valve *D* remains unaltered in position. Its action is also the same as in the suction pump. During the upstroke the valve *D* opens to let in water into the cylinder, the valve *G* remaining closed. It is the downstroke which forces water into *H* through *G*. It is subject to the same limitations as the force and the suction pumps.

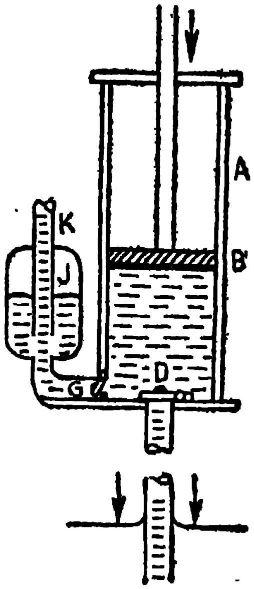


Fig. 56

The continuous action force pump. By adding an air-chamber (*J*, Fig. 56) to the force pump, the outflow of water may be made continuous. The action of the pump keeps the air in *J* compressed. The compressed air forces the water up the delivery tube *K* and maintains a continuous outflow.

54. Bicycle Pump

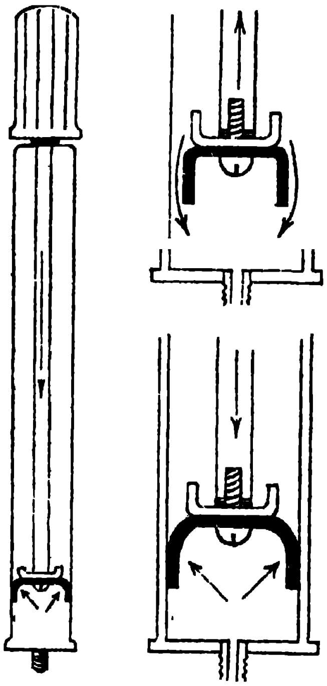


Fig. 57

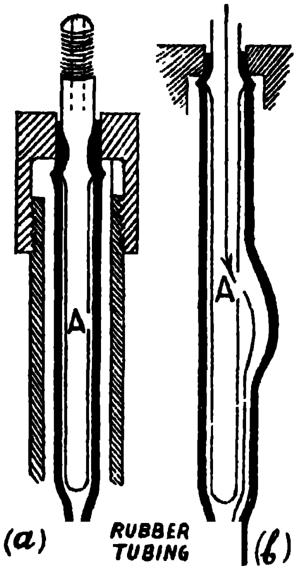


Fig. 58

The bicycle pump is a *compression pump*, the action of which may be understood from figures 57 and 58. The cup shaped leather washer attached to the piston head acts as a valve. At the upstroke air passes between the washer and the wall of the cylinder. At the downstroke the air is compressed. The compressed air presses the sides of the washer against the wall and prevents leakage. When air in the cylinder is sufficiently compressed, it forces open the valve in the tube (Fig. 58). This valve consists of a narrow metal tube with a small hole in the side covered with a thin rubber tube. The arrangement allows flow of air only in one direction.

A bicycle pump with a reversed washer and valve will act as a suction pump.

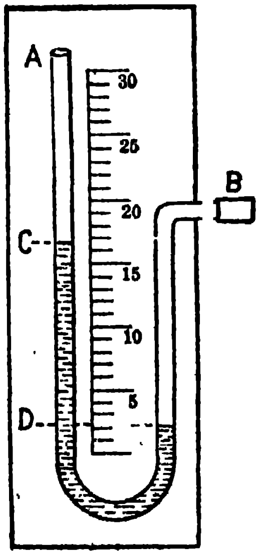


Fig. 59

55. Pressure Gauges

While the barometer is used for measuring the pressure of the atmosphere, pressure gauges (or manometers) are used for measuring the pressure of confined gases. We shall consider (i) the open-tube manometer, (ii) the closed-tube manometer and (iii) the Bourdon gauge for measuring pressure.

(i) **Open-tube manometer.** When the pressure does not far exceed atmospheric pressure, as the pressure of a domestic gas supply, the open-tube manometer (Fig. 59) may be used with profit. It consists of a U-tube open at both ends and contains a liquid which may be oil or (coloured) water when the difference of pressure with the atmosphere is rather small. For larger differences mercury is used. The way it may be connected to a gas tap to read the pressure of domestic gas supply is obvious from fig. 59. The scale attached to the manometer reads the pressure difference with the atmosphere, and may be expressed as so many inches of water or mercury above or below the atmosphere.

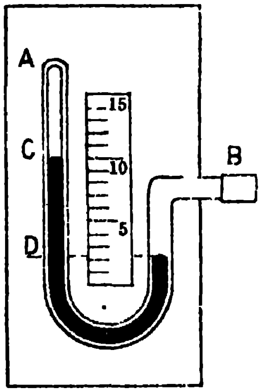


Fig. 60

(ii) **Closed-tube manometer.** If the pressure of the gas is much less than atmospheric pressure, a closed U-tube manometer may be used to measure the pressure. One limb of the U-tube is closed (Fig. 60). Mercury completely fills the closed limb. The open limb is connected with the vessel which contains the gas. If $CD = h$ is the difference in level of the liquid in the two limbs, the gas pressure is equal to the hydrostatic pressure of this liquid column. For very small pressures an oil may be used for mercury.

(iii) **Bourdon Gauge**. A more convenient type of pressure gauge is the Bourdon gauge. It can be adapted to measure pressures above or below atmospheric pressure, and is suitable for many purposes such as measuring the steam pressure inside the boiler of a steam engine or the pressure of compressed air in a cylinder.

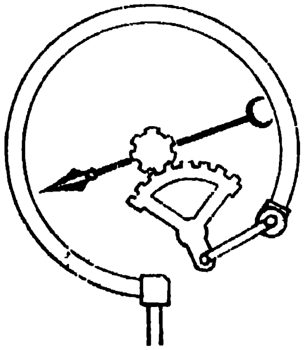


Fig. 61

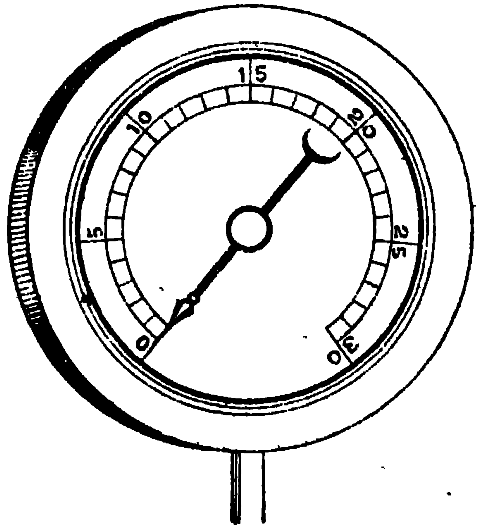


Fig. 62

The Bourdon gauge consists of a bronze tube of elliptical cross section, bent into a nearly complete ring and closed at one end (Fig. 61). The open end of the tube is connected with the chamber in which the gas pressure is to be measured. The closed end is connected by levers and a curved rack and pinion to a pointer which moves over a graduated scale. As the pressure in the flat tube increases, it tends to straighten out. This makes the pointer move over the scale. The scale is generally graduated in pounds per square inch, and reads zero when the pressure is one atmosphere. A pressure of 10 lb. *by the gauge* is, therefore, a pressure of 10 lb. per sq. in. above one atmosphere.

In Bourdon gauges meant for measuring pressures less than one atmosphere, the scale is marked from 0 to 30 (Fig. 62). The graduations are in inches of mercury. A reading of 30 would mean a nearly perfect vacuum. A reading of 10 would mean that 10/30 of the air has been removed.

Exercises

1. Explain how the siphon acts. Under what conditions will it fail to act ?

2. How does a common pump, as may be fitted to a tube well, work ? Can the pump draw water from any depth ? Explain your answer.

What is *priming* ? Why is it necessary ?

3. A tube well at ground level may draw underground water from a depth of 200 ft. A manually operated pump may raise the water to a height of 40 ft. above the ground.

How can the above be reconciled with the fact that atmospheric pressure cannot support a column of water taller than 34 ft. ?

4. How would you arrange for a continuous delivery of water from a water-pump ?

5. Explain how a bicycle pump acts as an air compressor. What part is played by the cup shaped leather washer ?

6. How would you measure the pressure of the gas supply in your laboratory ? How would you express the result of your observation ?

7. Draw a diagram of a Bourdon gauge and say how it acts. What are the actual pressures when the gauge readings are (i) 5 atmospheres, (ii) 20 inches of vacuum ?

MOTION AND FORCE

56. Elementary Ideas

Mechanics. The science of mechanics is the most fundamental of all the branches of physics. It deals with such ideas as motion, force and energy, and their relations to each other. A knowledge of mechanics is essential for the proper understanding of physical phenomena.

Mechanics has two broad divisions, viz., dynamics and statics. Dynamics deals with bodies in motion, and statics, with bodies at rest.

The Material Particle. In mechanics the term 'material particle' or, simply, 'particle' means a body of negligible size. It is represented by a point. We have to consider its position and mass only.

Rest and Motion. When we say that a body is at rest we ordinarily mean that it is at rest with respect to the surface of the earth. A body in motion means a body which is moving relative to the surface of the earth.

Displacement. A body is said to be displaced when its position changes relative to its surroundings. Let a body, here represented by a point, which was initially at O (Fig. 63), move to A. Whatever path the body follows in moving from O to A, its displacement is measured by the length and direction of OA. At some later time the body may move to B. Then it will be said to have a further displacement AB. The total displacement from the initial position will be OB.

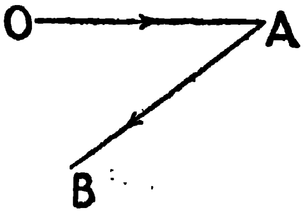


Fig. 63

Displacement always conveys the idea of direction as well as of magnitude.

57. Speed and Velocity

Speed. Speed of a body is the distance it moves in unit time. If it covers a distance s in time t its speed is s/t.

If, while moving, a body covers equal distances in equal times, its speed is said to be *uniform*. If it does not do so, the speed is *non-uniform* or *variable*. When speed is variable, s/t represents the average speed over the distance s or the time t.

Units. Speed is expressed in *units of length per unit of time*, such as cm. per sec., km. per hour, ft. per sec., miles per hour (or m.p.h. for brevity) etc. The *knot* is a nautical unit of speed and implies a speed of one sea-mile (i.e., 6020 ft.) per hour. It is correct to say 'a speed of 20 knots'.

Velocity. When a body moves in a definite direction, the speed of the body in that direction is called its velocity. Remember that *velocity always implies speed in a fixed direction.* While speed has magnitude only, velocity has both magnitude and direction.

Like speed, velocity may be uniform or variable and is expressed in the same units as speed, but with the direction mentioned, such as 'a velocity of 50 m.p.h. due north'. Velocity will change when the speed changes, or the direction, or both.

Velocity may also be defined as displacement per unit time.

Time-Velocity Graph. The velocities of a moving body at successive moments may be plotted graphically against the time and the points connected together by a smooth curve. Such a line is called a time-velocity graph. Let the velocities at different times be as follows :

Time in seconds.	0	1	2	3	4	5
Velocity in ft. per sec.	2	2·4	3	3·5	3·8	4

Plot the time along the x-axis and the velocity along the y-axis. Join the points so plotted by a smooth line (Fig. 64). This line is the time-velocity graph.

If the velocity is uniform, the graph will be a straight line parallel to the x-axis. *The distance traversed between any two moments will be given by the area bounded by the ordinates for the initial and final moments, the graph and the x-axis.* This statement holds even when the velocity changes with time.

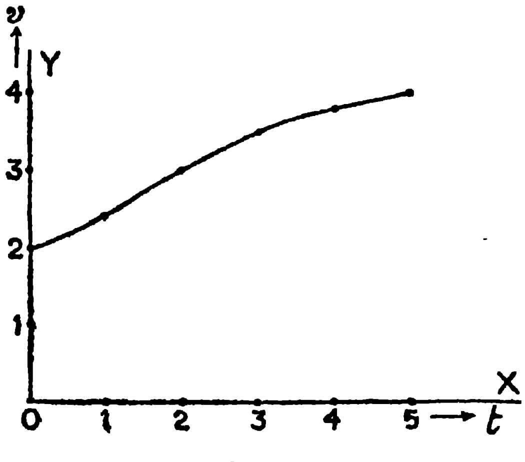


Fig. 64

58. Acceleration

When a train leaves a station it gradually gains in speed. Starting from rest it moves faster and faster till it acquires a practically constant speed. At the initial stages, when it is gaining in speed, it is said to be *accelerated.* When brakes are applied to bring it to a stop, it gradually loses speed and finally stops. It is then said to be *retarded* or *decelerated.* In both cases the speed changes with time.

A velocity will change if the speed changes. *Rate of change of velocity is called acceleration.* An acceleration is constant or uniform when the velocity changes by equal amounts in equal intervals of time, however small those intervals of time may be.

The following table illustrates a case of uniform acceleration.

Time in sec.	0	1	2	3	4	5	6
Velocity in ft. per sec.	2	2·5	3	3·5	4	4·5	5

The velocity changes by 0·5 ft. per sec. in every second. This is expressed by saying that the acceleration is 0·5 ft. *per second per second* or 0·5 ft. *per* $\textit{sec}^2$. In all expressions of acceleration the term 'per sec.' occurs twice. This is due to the fact that the unit of time is involved twice in such expressions,—once, in the velocity, and then again in the rate of change of velocity.

The time-velocity graph of the values in the above table is shown in fig. 65. It is the straight line *AB*. When the acceleration is constant, the time-velocity graph will be a straight line.

Deceleration or retardation is acceleration with a negative value. Here the velocity diminishes with time. If the initial velocity is 8 ft. per sec., and after 2 sec. it comes down to 2 ft. per sec., the deceleration is (8 ft. per sec.—2 ft. per sec.) /2 sec., i.e., 3 ft. per sec. per sec. The acceleration is – 3ft. per sec. per sec.

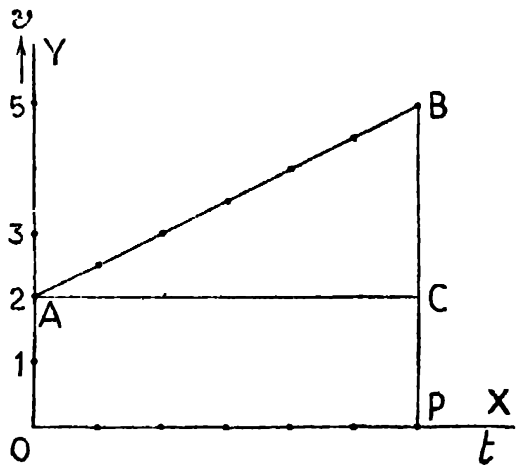


Fig. 65

59. Algebraic Formulæ for Uniformly Accelerated Motion

Let u = initial velocity of a body,
v = its final velocity,
t = the time in which this change occurs, and
f = the acceleration,

all quantities being expressed in the same system of units.

Since f is the increase in velocity per second, the velocity after one second is $u+f$; after two seconds, $u+2f$ etc. Hence at the end of the time t seconds, the velocity will increase by ft. The *of a body or its direction of motion,* or briefly, *a force can impart acceleration to a body.*

A useful concept in the relation between force and motion is **momentum.** *It is measured by the product of the mass of a body and its velocity.* Consider a cricket ball thrown softly; it will be easy to stop it. But when hit hard, it is much more difficult to stop. In the latter case it has a larger momentum.

Think of a heavy iron roller lying on the ground or a four-wheeled carriage standing on a pavement. Because of its large mass you find it difficult to move. A continued push applied for some time gives it a certain velocity. A harder push over a shorter period will give rise to the same velocity. So, to give the body a momentum you had to apply a force on it for a certain period. To bring the moving roller or the carriage to rest, i.e., to destroy the momentum the body possesses, you must oppose the motion with a force and continue to exert the force for some time until the body stops. Production or destruction of a larger momentum requires application of a larger force for the same time or of the same force for a longer time.

62. Other Effects of Force

What does a force do besides producing a change in the motion of a body? It can overcome other forces. The cases of friction and gravity are within the compass of one's constant experience. A body will not move unless we apply enough force to overcome the friction which opposes its motion. Even when moving, a body experiences a frictional opposition to motion. A part of the force applied to keep the body in motion goes to overcome this frictional opposition. To lift a body we must apply enough force to overcome gravity.

Forces may act on a body without producing motion. Such forces deform the body. When we squeeze a ball or apply a pull to a spring, we do not set it in motion, but deform it. In all cases where a body does not move though forces act on it, the body is deformed by the forces. The extent of the deformation depends on the magnitude of the forces and the nature of the body. The study of elastic properties of materials concerns itself with these effects (see § 79).

63. Newton's Laws of Motion

The exact relation between force and motion was set forth by Newton in the form of three laws known as Newton's laws of motion. These laws form the basis of mechanics and it is impossible to estimate how difficult the subject might have been without them. A formal proof of the laws is not possible, but the results obtained by applying these laws in various fields, are found to be in agreement with facts. The laws may be stated as follows :

Law I. *Every body continues in its state of rest or of uniform motion in a straight line unless impelled by an external force to change that state.*

Law II. *The change of momentum per unit time is proportional to the impressed force, and takes place in the direction in which the force acts.*

Law III. *To every action there is an equal and opposite reaction.*

64. Discussion of the First Law—Inertia·

It is common experience that a body at rest continues to be so unless an external agency makes it move. But we do not find a body in motion to continue to move uniformly in a straight line. It actually comes to rest sooner or later. The reason for the stoppage is that there is a force, the force of friction, opposing its motion. We can well imagine that if opposing forces were fully removed, a body would continue to move in a straight line with uniform velocity. For it is our experience that while a ball rolls a short distance on a rough floor, it traverses longer distances the smoother the floor.

The tendency of a material body to continue in its state of rest or of uniform motion in a straight line is a fundamental property of matter and is called **inertia.** Due to inertia matter resists any attempt to start it if at rest, to stop it if in motion, or in any way to change either the direction or the speed of motion. It is due to inertia that a body on a moving train tends to move towards the forward end when the train stops, and towards the rear end when the train starts. In both cases the body tends to continue in its previous state whether that was one of rest or of motion. Other examples of inertia of motion are the particles of mud which fly off tangentially from the wheel of a moving automobile, or the sparks that fly off in straight lines from a grinding wheel when a metal is pressed on to it. A ball thrown upwards in a moving train moves with the train because of its inertia of motion.

Definition of Force. The first law also defines force. It implies that an *unbalanced* force must be impressed on the body from outside to bring about a change in its state of rest or of motion.

If there are opposing forces, a motion may not be produced by a force. For example, we may push a heavy piece of furniture without moving it. Here forces of friction prevent motion. If these were absent, the external impressed force would have moved the piece of furniture. The push might have failed to produce motion, but undoubtedly it tended to do so.

We may, therefore, define force by saying that:

Force is the agency which changes or tends to change the state of rest or of motion of a body.

65. The Second Law—the Fundamental Kinetic Equation

While the first law defines force, the second law provides a means of measuring it.

Let u = initial velocity of a body,
v = its final velocity,
t = the time in which this change occurs,
m = its mass, and
f = its acceleration.

The change of momentum in time $t = mv - mu$.

$$\therefore \text{ the rate of change of momentum } = \frac{m(v-u)}{t} = mf.$$

$$= \text{mass} \times \text{acceleration}.$$

According to the second law the change of momentum, and hence the acceleration, will take place in the direction in which the force acts.

If P is the applied force, we have, from the second law

$$P \propto mf$$

$$\text{or} \quad P = kmf \tag{65·1}$$

where k is the factor of proportionality and is a constant. The value of this constant depends on our choice of units of P, m and f. If we agree that the force which produces unit acceleration on unit mass will be called the unit of force, then in Eq. 65·1, $P=1$, $m=1$ and $f=1$. This makes $k=1$. On this basis we write

$$P = mf \tag{65·2}$$

or Force = Mass × Acceleration

This is the *fundamental kinetic equation* and gives us a means of measuring force, as also mass. For according to Eq. 65·2 mass of a body may be defined as the ratio of the force to the acceleration it imparts to the body. The mass, so defined, is a measure of the inertia of the body.

Impulse of a force. The product of a force (P) and the time (t) for which it acts on a body is known as the impulse of the force. An impulse brings about a change of momentum. For

$$P \times t = m \times f \times t = m\ (v-u) \tag{65·3}$$

66. Units of Force

Absolute units. A unit of force defined according to the equation $P = mf$ is called an absolute unit. In the c.g.s. system the absolute unit of force is the dyne. *The* **dyne** *is that force which, acting on a mass of* 1 *gm, gives it an acceleration of* 1 *cm/sec*^{2}.

$$1 \text{ dyne} = 1 \text{ gm} \times 1 \ \frac{\text{cm}}{\text{sec}^2} = 1 \frac{\text{gm.cm.}}{\text{sec}^2}$$

$\left[\text{Dyne can be replaced by its equivalent } \frac{\text{gm. cm.}}{\text{sec}^2} \right]$

In the British system the absolute unit of force is called the poundal. *The* **poundal** *is that force which, acting on a mass of* 1 *lb., gives it an acceleration of* 1 *ft/sec*^{2}.

$\left[\text{Poundal can be replaced by its equivalent } \frac{\text{lb.ft.}}{\text{sec}^2.} \right]$.

Relation between the poundal and the dyne.

$$\begin{aligned} 1 \text{ Poundal} &= 1 \text{ lb.} \times 1 \text{ ft/sec.}^2 \\ &= 453{\cdot}6 \text{ gm} \times 30{\cdot}48 \text{ cm/sec.}^2 \\ &= 453{\cdot}6 \times 30{\cdot}48 \text{ gm. cm/sec.}^2 \\ &= 1{\cdot}382 \times 10^4 \text{ dynes.} \end{aligned}$$

Gravitational Units of Force. Much earlier in the text we have used such terms as 'a pound of force or a pound-weight' and 'a gram of force or a gram-weight'. These are gravitational units of force and represent the attraction of the earth on a unit mass, i.e., the weight of a pound or a gram. As the weight of a body varies slightly at different places it is necessary for the sake of accuracy to specify where this weight is measured.

A *pound of force* (or a pound-weight) is the weight of a pound of mass at 45° latitude and sea-level.

A *gram of force* (or a gram-weight) is the weight of a gram of mass at 45° latitude and sea-level.

For most practical purposes the slight variations in weight due to differences in location can be neglected.

Since a pound of force makes a pound of mass move with an acceleration g ($=32$ ft/sec^2), we have

$$\begin{aligned} 1 \text{ lb. wt.} &= 1 \text{ lb.} \times g \text{ ft/sec.}^2 \\ &= g. \text{ lb. ft/sec.}^2 \\ &= g \text{ poundals } (=32 \text{ poundals}). \end{aligned}$$

Similarly, 1 gm. wt. $=980$ dynes.

67. Third law—Action and Reaction

The terms action and reaction are simply other names for force. If a body A applies a force on another body B, which we call the *action*, then, according to the third law, the body *B* will apply an equal and opposite force on *A*, which is called the *reaction*. The law thus states that *in nature forces always occur in pairs*, and that *each force of the pair acts on a different body*. Whether the bodies are at rest or in motion, touch each other or are separated by a distance (as in the case of two magnets), the law is of universal application.

Consider a book lying on a table. The weight of the book exerts a downward force on the table. According to the third law, the table will exert an equal and opposite force on the book. If the former force is called the action, the latter force is the reaction. Note that *reaction lasts only so long as the action is there*. Besides, *action and reaction act on two different bodies, and cannot, therefore, produce equilibrium*. For equilibrium the forces on one and the *same* body must be balanced.

Examples. (i) When a ladder leans against a wall it presses upon the wall. This pressure, here called the action, tends to overturn the wall. The counter-thrust exerted by the wall on the ladder is the reaction. It keeps the ladder in position.

(ii) Consider a rope tied to a tree and a man pulling at the other end of the rope. The force exerted by the man on the rope is the action. The reaction is the equal and opposite force exerted by the rope on the man.

The force exerted by the man on the rope is transmitted through the rope to the tree. The rope thus exerts an action on the tree. The tree exerts an equal and opposite reaction on the rope. The rope is in equilibrium under the action of the pulls exerted on it by the man and the tree.

[In this example the rope transmits two forces through it—one exerted by the man which acts through the rope on the tree, and the other, the reaction to this force which is exerted by the tree on the man. The rope is stretched and is said to be in 'a state of tension'. *Either force* transmitted through the rope is called a *tension*. This would show that when two men pull at two ends of a string, each with a force of 100 lb., the tension in the string is still 100 lb.]

68. Vector and Scalar Quantities

In the course of our studies we have come across several quantities which, besides magnitude, have also a direction associated with them. The term *vector* is applied to quantities which have both direction and magnitude. Displacement, velocity, acceleration, momentum and force are examples of vector quantities. The term *scalar* refers to quantities which have magnitude only. Mass, volume, density, speed etc. are examples.

Any vector quantity may be represented by a straight line drawn in the appropriate direction with an arrow head to indicate the sense and having a length which represents, to some convenient scale, the numerical value of the quantity. A velocity of 30 mi/hr due north and another of 40 mi/hr due east may be represented by two straight lines drawn at right angles and having lengths in the ratio 3:4 (Fig. 67).

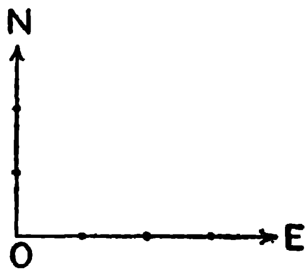


Fig. 67

In the case of a force, in addition to its magnitude and direction, we have to consider the point at which it is applied. *The point of application of a force may be represented by the point from which the straight line is drawn.*

Addition of Vectors. Scalar quantities are added algebraically; but not so the vectors. Vectors may be added geometrically as shown in Fig. 68. Let OA and OB be the vectors to be added. Complete the parallelogram $OACB$. Then OC, the diagonal, will represent in magnitude and direction the vector sum of OA and OB.

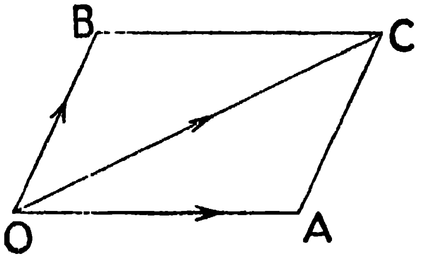


Fig. 68

Exercises

1. Distinguish between speed and velocity.

Draw a time-velocity graph from the following data and find the distance traversed :

Time in seconds	0	1	2	3	4	5	6
Velocity in ft/sec.	20	18	15	11	6	2	0

2. Define acceleration. Give two common examples each of acceleration and deceleration. Why does 'time' occur twice in an expression of acceleration ?

Express an acceleration of 1 cm per sec. per sec. in metres per min. per min. (*Ans*: 36)

[*Hint* : 1 cm. per sec. per sec $=1\frac{\text{cm}}{\text{sec}^2}=x\,\frac{\text{m}}{\text{min}^2}$

$\therefore\ x=1.\left(\frac{\text{min}}{\text{sec}}\right)^2\times\frac{\text{cm}}{\text{m}}=\ldots\ldots$]

3. (a) Deduce the formula $s = ut + \frac{1}{2}ft^2$ graphically.
(b) Derive the equation $v^2 - u^2 = 2fs$.

In both cases clearly state the meanings of the symbols used.

4. What do you understand by the statement that the acceleration due to gravity is 32 ft. per sec. per sec.?

5. It is stated that all bodies fall to the earth with the same acceleration. But when a piece of paper and a piece of stone are simultaneously dropped, the latter reaches the earth first. What is the reason ?

How would you show that the statement in the first sentence is true ? What is a 'free fall' under gravity ?

6. A body is projected vertically upwards. What will be its equations of motion ?

7. What will be the equations of motion of a body falling freely under gravity ?

8. What principal effects do forces produce ? Give an example of each kind.

9. State Newton's laws of motion, and comment briefly on each.

(*N.B.*—Carefully select the material you would include in the comments. The answer should be written in about 20 minutes).

10. Discuss how the concepts of *inertia* and *force* are contained in Newton's first law of motion.

11. Deduce the relation $P = mf$ from Newton's second law of motion.

Define the *dyne* and the *poundal*. Express the latter in terms of the former.

12. What is a pound of force ? A gram of force ? How is each related to the corresponding absolute unit of force ?

13. Two boys pull at the ends of a string, each with a force of 50 lb. Apply Newton's third law of motion to the case and say what is the tension in the string.

14. What is a vector ? Name the vector quantities known to you.

How can a force be represented graphically in all its aspects by a straight line ? How can two forces acting at the same point be added ?

15. A body moves with a constant acceleration of 5 cm. per sec. per sec. If it had an initial velocity of 100 cm/sec, what will

be its velocity after 20 seconds? How far will the body move in that time?

(*Hint*: Use formulæ $v=u+ft$ and $s=ut+\frac{1}{2}ft^2$)

(*Ans*: 200 cm/sec.; 3000 cm.)

16. A car moving with a velocity of 30 mi/hr is decelerated by 4 ft/sec in every second. How long will it take to come to a stop and how far will it move in that time?

If you want the car to stop in (i) one-fourth the distance, (ii) one-fourth the time, what should be the deceleration in each case?

(*Hint*: $u-ft=v=0$; $s=ut-\frac{1}{2}ft^2$)

(*Ans*: 11 sec; 242 ft; 16 ft/sec^2; 16 ft/sec^2.)

17. A body falls freely from rest under gravity. Find how far it falls (i) in three seconds, (ii) in the third second.

(Take $g=980$ cm/sec^2). (*Ans*: 4410 cm; 2450 cm.)

18. A body is projected vertically upwards with a velocity of 80 ft. per second. Find how long it will rise. What maximum height will it reach? What will be its velocity when it is half-way up? ($g=32$ ft./sec.^{2})

(*Hint*: $v=u-gt=0$ at the top. $s=ut-\frac{1}{2}gt^2$; $v^2-u^2=-2gs$).

(*Ans*: 2.5 sec.; 100 ft; $40\sqrt{2}$ ft per sec.)

19. With what initial velocity must a body be projected vertically upwards so as just to reach a height of 144 ft.? When will it be 80 ft. above the ground? Explain the two values of the time. ($g=32$ ft./sec.^{2})

(*Ans*: 96 ft. per sec.; 1 sec. and 5 sec.; going up, going down.)

20. A force of 20 dynes acts on a body of mass 5 gm. What is the acceleration produced? If the body was initially at rest and the force acted on it for 5 seconds, how far will the body move in this time? (*Hint*: Use formulæ $P=mf$ and $s=\frac{1}{2}ft^2$.)

(*Ans*: 4 cm/sec.^{2}; 50 cm.)

21. The speed of a locomotive weighing 300 tons changes from 10 mi/hr to 40 mi/hr in 14 seconds. Find the force exerted in poundals. (*Ans*: 2.112×10^6).

22. A cricket ball weighing $5\frac{1}{2}$ oz. and moving at 60 ft/sec. is caught by a fieldsman who draws back his hands so as to bring the ball to rest in $\frac{1}{4}$ sec. What is the average force the ball exerts on his hands? [*Hint*: $Pt=m(v-u)$.] (*Ans*: 82.5 poundals.)

23. A motor-car travelling at 30 mi/hr approaches a traffic signal 440 yards away and has to stop there. If the car weighs 15 cwt. and is uniformly decelerated to a stop, find (i) the deceleration, (ii) the time which elapses before it comes to rest, and (iii) the braking force applied.

(*Ans*: 0.73 ft./sec.^{2}; 60 sec.; 1232 poundals.)

CHAPTER 8

WORK, POWER AND ENERGY

69. Work

In our everyday language the term 'work' means some kind of labour, physical or mental, producing a result. In the language of mechanics, or of science in general, the term is used in a special sense. *Work is said to be done when the point of application of a force moves in the direction of the force.*

Work, so defined, involves two quantities, viz., (i) a force and (ii) a distance in the direction of the force. When a man pushes a cart or lifts a load from the floor, he not only exerts a force, but exerts it through a distance in the direction of the force. Under these conditions the man is said to do work. When, however, he supports a bucketful of water in his hand without moving it, he exerts a force but does not exert it through any distance in the direction of the force. Whatever muscular fatigue he may experience in the act of holding the bucket, he does ***no work*** in the technical sense of the term.

The amount of work (W) done is measured by the product of the force (P) and the distance (s) through which the point of application of the force moves in the direction of the force. In symbols,

$$W = Ps$$

or **Work = Force × Distance**

It may so happen that the displacement of a body is not in the direction of the force under consideration. For example, when a body moves down an inclined plane under gravity, the force of gravity and the motion it produces are not in the same direction. In such a case work is measured by

$$W = Ps \cos \theta$$

where θ is the angle between P and s. It may be noted that $P \cos \theta$ is the effective part (component) of the force in the direction of s. Hence in such cases,

Work = Effective force × Distance.

***Work done* by, on *or* against.** When a body A exerts a force P on another body B making it move through a distance in the direction of P, we say

(i) A (the agent which applies the force) does work *on* B.

(ii) Work is done *by* A or the force P *on* B.

When an agent moves a body against an opposing force, work is said to be done *against* the opposing force. Thus when a man lifts a load from the floor, he does work *on* the load *against* gravity. When a body falls under gravity, work is done *by* gravity.

Units of Work. The unit of work involves the units of force and distance. Force may be expressed in absolute or gravitational units (§ 66). Besides, there are the metric and the British systems. All these lead to a multitude of units of work. The more important of these units are listed below.

Nature of unit	*Unit of force*	*Unit of distance*	*Corresponding unit of work*
CGS absolute	dyne	cm.	1 dyne × 1 cm =1 **erg**
,, gravitational	gm. wt.	cm.	1 gm. wt. × 1 cm =1 **gram-centimetre** (gm.cm)
FPS absolute	poundal	ft.	1 poundal × 1 ft. =1 **ft-poundal**
,, gravitational	lb. wt.	ft.	1 lb.wt × 1 ft. =1 **foot-pound** (ft. lb.)

A definition in words may be easily provided for each unit of work. Illustrating with the erg, we may say that the *erg* is the work done when the point of application of a force of one *dyne* moves through a distance of one *centimetre* in the direction of the force.

For other units appropriate words may be substituted in the above statement.

The **Joule**: The erg is a very small unit. The Indian Naya Paisa weighs 1·5 gm. Its weight is, therefore, $1{\cdot}5 \times 980 = 1470$ dynes. When lifted through a distance of one metre the work done $= 1470$ dynes $\times$ 100 cm $=$ 1,47,000 ergs.

The Joule is the practical unit of work in the c.g.s. system, and is equal to 10^7 ergs.

1 Joule $= 10^7$ ergs.

It is not difficult to understand the meanings of such terms as 'a kilogram-metre' or 'a ton-foot'. The former is the work done when a weight of one kilogram is lifted through one metre against gravity.

To express the foot-pound in joules.

$$1 \text{ ft. lb.} = 1 \text{ lb. wt.} \times 1 \text{ ft.} = 32{\cdot}2 \text{ poundals} \times 1 \text{ ft.}$$

$$= 32{\cdot}2 \text{ lb.} \frac{\text{ft.}}{\text{sec.}^2} \times 1 \text{ ft.}$$

$$= 32{\cdot}2 \times 453{\cdot}6 \times 30{\cdot}48 \ \frac{\text{gm. cm}}{\text{sec.}^2} \times 30{\cdot}48 \text{ cm.}$$

$$= 1{\cdot}356 \times 10^7 \text{ dynes} \times \text{cm.} = 1{\cdot}356 \text{ joules.}$$

70. Power

When work is done, its amount is not the only item of importance. The time in which the work is done is also of great importance. Suppose a 100 gallon tank is to be filled by drawing water from a well, 50 ft deep. 100 gallons of water weigh 1000 lb. Hence the work to be done in drawing it from the well will be $1000 \times 50 = 50{,}000$ ft. lb. If the work is done by manual labour a stronger man will be able to do it more quickly than a weaker one. The former would then have done more *work per unit time* than the latter. If a motor-driven pump were employed, a more powerful motor would complete the job in a shorter time than a less powerful motor. The former does more work per unit time than the latter.

The time rate of doing work is called power, i.e., *power is the work done per unit time.*

$$\textbf{Power} = \frac{\textbf{Work}}{\textbf{Time}} = \frac{\textbf{Force} \times \textbf{Distance}}{\textbf{Time}} = \textbf{Force} \times \textbf{Velocity}.$$

When a person does work his sense of fatigue is determined more by the rate at which he does the work than by the total work done. Suppose a boy weighing 120 lb. ascends a flight of steps which carries him 15 ft. above the ground. The work he does against gravity is $15 \times 120 = 1800$ ft. lb. When he completes the ascent in 20 seconds he feels little fatigue. But if he tries to do it in 4 seconds, he will feel much more fatigued. In the latter case his rate of doing work, i.e., power, is 5 times greater than that in the former case.

Two railway trains of the same weight move at 30 m.p.h. and 50 m.p.h. respectively. If the friction of the rails and the resistance due to air are the same for both, the latter does more work than the former, because it moves a greater distance in the same time. So its rate of doing work, or power, is greater than that of the former.

The power delivered by a machine may be constant, or it may fluctuate. In the latter case we shall be concerned with the *average power.*

Units of Power—Watt ; Horse-power. Any of the units of work combined with a suitable unit of time will provide a unit of power. Two practical units of power are, however, of great importance. They are (i) the watt and (ii) the horse-power.

When one joule of work is done per second, the power is one *watt*. A power of thousand watts is a *kilowatt* (kw.). In other words, when 1000 joules of work are done per second, the power is one kilowatt. Electric machines are rated in watts and kilowatts.

The practical unit of power used in engineering practice is the horse-power (h.p.). When work is done at the rate of 550 foot-pounds per second (or 33,000 ft. lb. per minute), the power is called one *horse-power*.

To express the horse-power in watts.

$$\begin{aligned} 1 \text{ Horse-power} &= 550 \text{ foot-pounds/sec.} \\ &= 550 \times 1{\cdot}356 \text{ joules/sec. (vide § 69)} \\ &= 746 \text{ watts.} \end{aligned}$$

A h.p. is thus approximately $\frac{3}{4}$ kw.

The kilowatt and the horse-power have given rise to two more *units of work*—the **kilowatt-hour** and the **horse-power-hour.** The kilowatt-hour is the work done when power is used at the rate of one kilowatt over a period of one hour. This is the unit on which the cost of electricity is based.

Similarly, the horse-power-hour is the work done when power is used at the rate of one horse-power over a period of one hour.

71. Energy

When a body can do work, we say it possesses energy. *Energy of a body is defined as its capacity for doing work.* It is measured by the amount of work the body can do. Energy and work are essentially the same kind of quantity, and are measured in the same units.

A man or a horse can do work ; so he possesses energy. Steam can push the piston within the cylinder of a steam engine ; so it possesses energy. A moving body possesses energy since it can make other bodies move when it collides with them, An elevated body possesses energy since, while falling, it can pull other bodies up. If such a body pulled a mass of 1 kg. through a height of 1 m. before reaching the ground, we may say its energy was 1 kilogram-metre or 10^5 ergs, or that the body lost 10^5 ergs of energy in pulling the weight up.

Different forms of energy. There are many forms of energy. An electrically charged body attracts other bodies, and can make them move. It possesses what we call electrical energy. Magnets possess magnetic energy. Heat of steam makes locomotives move. Heat is a form of energy. Radio waves, light and X-rays are also forms of energy. An important form of energy is chemical energy, as is possessed by coal. When coal burns it combines with oxygen releasing chemical energy in the form of heat.

In recent times man has harnessed a vast source of energy. It is derived from the conversion of matter into energy under the action of nuclear forces in the atom, and is called *atomic energy*. The atom bomb, the hydrogen bomb and the atomic pile are devices for the release of atomic energy.

In mechanics, we are however interested in what we call *mechanical energy*. A body may possess mechanical energy due to either one or both of two causes, viz., (i) by virtue of its *motion*, and (ii) by virtue of its *position* or *configuration*. The former is called *kinetic energy* and the latter, *potential energy*.

Kinetic energy of a body is defined as the energy it possesses by virtue of its motion. It is measured by the amount of work a moving body can do against an opposing force before it comes to rest. A moving hammer possesses kinetic energy, which enables it to do work in driving a nail into a wall against the resistance.

To obtain an expression for the kinetic energy of a body of mass m moving with velocity v, let us impress on the body a force P which opposes the motion and finally brings the body to rest. The acceleration (or rather the deceleration) produced by the force is $-f=-P/m$. If the body moves through a distance s before it comes to rest, then from the relation that

$$(\text{final velocity})^2-(\text{initial velocity})^2 = 2\times\text{acceleration}\times\text{distance} \quad (\text{Eq. } 59\cdot3)$$

we have

$$0^2-v^2=-2fs$$

$$\text{or} \quad v^2 = 2\frac{P}{m}s$$

$$\text{or} \quad Ps = \tfrac{1}{2}mv^2$$

Now, Ps is the work done by the moving body against the opposing force before coming to rest. Hence this is also the magnitude of its kinetic energy. Since $Ps=\frac{1}{2}mv^2$, the latter expression gives the kinetic energy of the moving body in terms of quantities which are known.

Kinetic Energy = $\frac{1}{2}$ mass × (velocity)^2

$$\mathbf{W=\tfrac{1}{2}\,mv^2}$$

It is easy to see that if a force P acting through a distance s changed the velocity of a body from u to v,

$$Ps=\tfrac{1}{2}m(v^2-u^2)$$

or *work done by the force = change in kinetic energy.*

Potential Energy is the energy which a body possesses by virtue of its position relative to the surroundings, its condition or configuration (i.e. the relative position of its parts). It is measured by the work the body can do in passing from the given position, condition or configuration, to some standard position, condition or configuration.

Consider a body of mass m raised to a height h above the ground. In falling it can pull another body up. The work it can do is given by the product of its weight and height above the ground (generally the standard position). This is the measure of its potential energy.

$$\textit{Potential energy} = mgh.$$

Compressed air can drive machinery. It possesses energy by virtue of its condition of being compressed. The standard condition is generally that under normal atmospheric pressure. The work that the compressed air can do in expanding to a pressure of one atmosphere is a measure of its potential energy.

The main spring of a clock, when wound, drives the hands of the clock. It acquires potential energy in winding which changes the relative position of the different portions of the spring. As it is unwound it loses energy.

The sum of the kinetic and potential energies of a freely falling body remains constant throughout its motion. Let m be the mass of a body falling freely from rest from a height h above the ground. Let v be its velocity after it has travelled a distance x from its initial position. From the relation

$$(\text{final velocity})^2 - (\text{initial velocity})^2 = 2 \times \text{acceleration} \times \text{distance}$$

we have

$$v^2 = 2gx$$

$\therefore$ The kinetic energy of the body at this height is

$$\tfrac{1}{2} mv^2 = mgx.$$

Its potential energy at this height, i.e., $h-x$ above the ground, is $mg\ (h-x)$.

$\therefore$ The total energy = Kinetic energy + Potential energy

$$= mg + mg\ (h-x) = mgh.$$

Since x may have any value from 0 to h, we find that the total energy of the falling body remains constant and is *equal to its potential energy at the topmost position.*

It is also clear from the analysis that *the gain in kinetic energy is equal to the loss in potential energy.*

72. Transformation of Energy

The case of a falling body is a very simple example of transformation of energy from one form into another. Here potential energy changes into the kinetic form. A body projected upwards shows the reverse transformation. When we look carefully at natural events, we find that they all provide illustrations of transformation of energy. In fact, a change is nothing but such a transformation.

When we rub our hands together heat is developed. Here mechanical energy is converted into heat. A steam engine converts heat energy obtained by burning coal into mechanical energy.

When a closed coil of wire is rotated by mechanical means in the neighbourhood of a strong magnet, electricity flows through the coil. Mechanical energy has been converted in the process into the electrical form. When an electric current passes through an electric motor, as in our electric fans or in the motors of tram cars or of the electric trains, the spinning of the motor illustrates transformation of electrical into mechanical energy. Examples of this kind may be multiplied without limit.

It is, however, interesting to note that *the sun is the ultimate source of energy for all changes on the earth.* Light and heat energy from the sun are necessary for the germination and growth of plants. Animals depend for their food on plants. Coal, one of our principal sources of energy, represents stored up sunshine of geologic ages. Oil, another important source of energy, is supposed to have been derived from the bodies of minute organisms which lived ages ago. The water-cycle, essential for life on earth, is maintained by the sun's heat.

73. Conservation of Energy

In all examples of transformation of energy we find that *a body cannot acquire energy except at the expense of energy possessed by other bodies. Energy can neither be created nor destroyed.* It can only change form, or pass from one body into another. Whenever one system loses energy another gains it. Where measurements are possible, it has been found that the loss of energy of one system is equal to the gain in energy of the other. When energy changes form a fixed quantity of one kind is always obtained for a fixed quantity of the other.

All these lead us to believe that the total amount of energy in the universe remains constant. This is known as the *law of conservation of energy.*

A system of bodies which can exchange energy only among themselves, no energy leaving the system nor any energy entering it from outside, is called an *isolated system.* The principle of conservation of energy may be stated in the form that *the total energy of an isolated system remains constant.*

Another important law of energy, of which we find innumerable examples and make simple use, is that *the potential energy of a system tends to become a minimum.* Thus bodies tend to roll down a slope and springs unwind.

74. Dissipation of Energy

Whenever energy is transformed, though none is destroyed, there is a loss of one kind. To run a machine energy is fed into it, and it does work. The effective work it does is always less than the work supplied to it. In other words, the *output is always less than the input.* Part of the energy is spent in overcoming the friction of moving parts, and appears as heat. The loss is not in the total energy, but in the availability or usefulness of some of it. Lord Kelvin was the first to recognise the general principle that whenever energy is used or transformed, some of it leaks away out of our control and becomes forever dissipated or unavailable. This principle is known as the *principle of dissipation of energy.*

Exercises

1. Define work. Distinguish between work done *by* a force and work done *against* a force giving two examples of each.

2. Name the common absolute and gravitational units of work and define any two of them. Say also how the two you define are related to each other.

3. What is meant by power ? Define the kilowatt and the horse-power and express one in terms of the other.

4. What is energy ? Distinguish between kinetic and potential energy, giving examples.

Show that the total energy of a freely falling body remains constant.

5. What do you understand by the principle of conservation of energy ? Give six examples of transformation of energy of different kinds. Which of them have large industrial application ?

6. What do you understand by dissipation of energy ? How does it affect machines which supply power ?

7. A pendulum bob is attached to a thread 100 cm. long. It is

pulled aside so that the thread makes an angle of 10° with the vertical and is then released. Find the velocity of the bob as it passes the lowest point.

Hint : Max. pot. energy of bob=its max. kinetic energy. Pot. energy$=mgh=mg\times100\ (1-\cos 10°)$. (*Ans* : 54·6 cm/sec).

8. A bullet of mass $\frac{1}{2}$ oz. has a velocity of 2400 ft./sec. Find its kinetic energy and the average resistance in lb. wt. offered by a body which it penetrates to a depth of 6 in.
(*Ans* : 90,000 ft.pdls ; 5625 lb. wt.)

9. A pump raises 30 gallons of water a minute through a height of 70 ft. What is the increase per minute in the potential energy of water ? What is the horse-power of the engine required to operate the pump if the efficiency of the engine is 80% ?

[Hint : Efficiency is the ratio of the output to the input.
(*Ans* : 21000 ft. lb. per min. ; 0·795 H.P.)

10. How fast can a crane lift a block of iron weighing 1650 lb. if 4 H.P. be applied ? (*Ans* : $\frac{4}{3}$ ft./sec.).

CHAPTER 9

PROPERTIES OF MATTER

75. Three States of Matter

A casual glance at our surroundings tells us that matter can exist in three states, viz., (i) solid, (ii) liquid, and (iii) gaseous. For purposes of simple classification we may say that—

A *solid* has a definite shape and a definite volume.

A *liquid* has a definite volume, but no definite shape. It assumes the shape of the vessel in which it is put.

A *gas* has neither a definite shape nor a definite volume. It completely fills the vessel holding it.

Since both liquids and gases can flow they are included together under the name **fluids.**

The three states in which we may find a substance such as water, are due to the differences in the states of aggregation of its molecules. In solids, molecules are arranged in regular patterns and cannot leave their places in the pattern. In liquids, molecules form closely packed swarms, but a molecule is almost free to move within a swarm. In gases, molecules are free to move about. Hence a gas completely fills the vessel holding it.

76. General Properties of Matter

Matter has some general properties which do not depend on the state of aggregation of its molecules. Some of the more important of these properties are :

(i) **Inertia.** Matter cannot of itself change its state of rest or of motion. This property has been discussed earlier in § 64.

(ii) **Gravitation.** Every particle of matter attracts every other particle in the universe with a force which is inversely proportional to the square of the distance between them and directly proportional of their masses (see § 77).

(iii) **Elasticity.** Matter resists any change of shape or volume. When the deforming force is removed it tends to come back to its original condition. This property of matter is called elasticity.

(iv) **Cohesion and Adhesion.** Molecules of the same kind attract each other strongly at very short distances. This is known as cohesion. Between molecules of different kinds there is a similar force of attraction, strong at very short distances. This is known as adhesion. These should not be confused with gravitational attraction.

Cohesive forces between glass molecules maintain the shape of a piece of glass and oppose its being broken into pieces. When water adheres to glass it does so on account of adhesion between water and glass molecules.

77. Gravitation and Gravity

The law of universal gravitation, first enunciated by Newton, states that.—

Every particle of matter in the universe attracts every other particle with a force F which varies directly as the product of their masses (m,m') and inversely as the square of the distance between them. In symbols, we may write

$$F \propto \frac{mm'}{r^2} \text{ or } F=G\frac{mm'}{r^2}$$

G in this equation is called the universal gravitational constant. In the c.g.s. system its value is $6{\cdot}67\times10^{-8}$. This means that two point particles of mass 1gm. each placed at a distance of 1 cm. from each other attract each other with a force of $6{\cdot}67\times10^{-8}$ dyne.

Spheres, solid or hollow, attract each other as if the entire mass of each were concentrated at its geometrical centre.

In accordance with the law of universal gravitation the earth attracts all bodies near it with a force proportional to the mass m of the body. Taking the earth to be a homogeneous sphere of radius R and mass M, the gravitational attraction, W, which it exerts on a body on its surface, is given by

$$W=G\frac{Mm}{R^2}$$

This force, called the **force of gravity** is directed towards the centre of the earth. *This is also the weight of the body.* We know that when a body falls freely under gravity it does so with a constant acceleration which we denote by g (§ 60). Hence by Newton's second law of motion the weight W of a body of mass m is given by $W=mg$.

$$\therefore \quad mg=GMm/R^2$$
$$\text{or} \quad g=GM/R^2.$$

Variation of g. We have stated in § 60 that g varies from place to place on the surface of the earth as the earth is neither a true sphere nor is it homogeneous. g increases in value with latitude. It has the lowest value at the equator and the highest value at the poles. Besides, g diminishes away from the surface of the earth, both above and below. At the centre of the earth $g=0$,

Mass and Weight. Since $W=mg$, and g varies in the manner just stated, it will be clear that the weight of a body will change (though very slightly) as we take it from one place to another on the surface of the earth. The weight will diminish if the body is moved to places above or below the surface of the earth. The changes are, however, very small. It should be noted that the mass of the body does not change in any way by such transfers.

Since the extension of a spring balance is determined by the weight of the body suspended, a spring balance is theoretically capable of detecting the changes in weight of a body. The graduations of such a balance, made at one place, will not strictly be correct at another place where g differs.

As, however, the differences are very small (about 1 in 400 at most) such discussions have more of an academic interest in them than any practical one.

78. The Simple Pendulum—Measurement of g

In view of the importance of the quantity g, the acceleration due to gravity, it is natural to enquire if there are simple methods of measuring this quantity. There are many, the one using a simple pendulum being among the easiest. Before describing the method it may be useful to add a few words about the simple pendulum.

A simple pendulum, strictly speaking, is an idealized device, being supposed to consist of a heavy particle suspended from a fixed point by a weightless, inextensible and perfectly flexible string. Obviously no actual pendulum can correspond to the above ; but a heavy spherical bob suspended by means of a light string is a fairly close approximation.

Laws of simple pendulum. The motion of a simple pendulum has certain characters which are often referred to as the 'laws of simple pendulum.' They are as follows :

(*i*) *Law of isochronism.* At a given place, the oscillations of a simple pendulum of given length are executed in equal times provided the arc of swing (i.e., angular amplitude) is small.

(*ii*) *Law of length.* For pendulums of small amplitude at a given place, the periodic time, T, is proportional to the square root of the length, l.

(*iii*) *Law of gravity.* For a given pendulum of small amplitude, the periodic time varies inversely as the squre root of the acceleration due to gravity, g.

(*iv*) *Law of mass.* The periodic time of a pendulum does not depend on the mass or the material of the bob.

Experiments confirming laws (*i*) (*ii*) and (*iv*) have already been described in § 21.

It will be worth while to remember that the periodic time of a simple pendulum of small amplitude is given by

$$T=2\pi\sqrt{\frac{l}{g}} \qquad (78.1)$$

Because g has the largest value on the surface of the earth, T will be shorter on the surface than anywhere above or below it. When T is shorter we say the pendulum is faster. As l shortens with fall in temperature, T will be shorter in winter than summer and the pendulum faster.

Measurement of g. Rearranging Eq. 78·1 we may write

$$g=4\pi^2\frac{l}{T^2} \qquad (78.2)$$

If we follow the method described in § 21 for Experiment 11 we can get the value of l/T^2 for different values of l. The mean value of this quantity as obtained from the experiments may then be substituted in Eq. 78·2 to give us the value of g. If l is in centimetres and T in seconds, g will be in cm. per sec. per sec.

(Question 4 of the 'Practical Exercises' at the end of Chapter 1 may be seen in this connection.)

Simple Harmonic motion. The motion of the bob of a simple pendulum oscillating with small amplitude is of special interest. If you study it you will see that at the end of a swing the bob is momentarily at rest. As it moves towards its normal position of rest it moves increasingly faster. The motion is fastest at the bottom of the swing. Away from this position the bob slows down until it comes to a momentary halt at the end of the swing. The same motions are then repeated.

The motion of the bob is one of variable acceleration. The most important character of the motion is that the acceleration is always directed towards the normal position of rest and, at any moment, it is proportional to the distance of the bob from this position.

Since the force acting on a body is equal to the product of its mass and acceleration, the force acting on the bob of an oscillating pendulum has all the above characters. A motion with these characters is called simple harmonic motion. We may lay down the following definition :

If a body moves under the action of a force which is constantly directed towards a fixed point and is proportional to the distance of the body from this fixed point, the motion is said to be simple harmonic.

79. Elasticity—Hooke's Law—Young's Modulus

We have stated in § 62 that while unbalanced forces produce motion, *balanced forces acting on a body produce deformation*. Consider a spiral spring (Fig. 69) supported rigidly at its upper end and weighted at the lower. If the lower end is pulled downwards the spring is acted on by two forces which balance, viz., (i) the applied downward pull, and (ii) an equal and opposite pull exerted by the support on the spring. The result of these two forces is an elongation—a deformation—of the spring. Very often we leave out of our consideration the force exerted on the spring by the support and speak of the elongation as being due to the pull exerted on it.

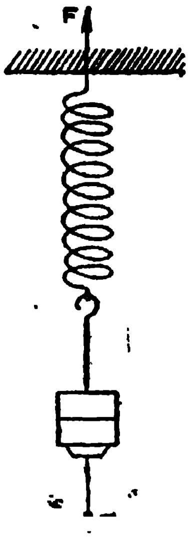


Fig. 69

In the above example the hand that pulls the spring experiences an upward force. This upward force resists the deformation of the spring, and is larger the greater the applied pull. When the pull is removed, the spring acquires its original length. This simple observation shows that—

(i) When a body (here the spring) is deformed, the body resists the deformation, and

(ii) When the deforming force is removed, it tends to go back to its original condition.

Elasticity. The property of matter by virtue of which matter resists the action of a force tending to change its shape or size and returns to its original shape or size on removal of the deforming force, is called **elasticity.**

Strain. If the shape or size of a body changes under the action of a force, the body is said to be *strained*. The change in shape or size that it undergoes is its *deformation*. The fractional deformation or the relative change so produced is called its *strain*.

Stress. When a body is strained, forces of reaction are brought into play within the body. These forces, which resist the deformation and tend to restore the body to its original form and dimensions, constitute a **stress.**

Units of strain and stress. Consider a piece of wire, rigidly supported at its upper end, being stretched by a load at its lower end. As the load increases, so does the elongation. If

e = the elongation for a load F,
l = the initial length of the wire,
$\propto$ = area of cross section of the wire,

then strain $= e/l$ and stress $= F/\propto$. Strain, being a ratio of two similar quantities, is expressed as a *pure number*. F is in reality a force. Hence stress is to be expressed as a *force per unit area*. Note that this is also the way we express a pressure. Often F is expressed in mass units.

Elastic limit. In the above example of the stretched wire, we find that the wire resumes its initial length on removal of the load provided the load does not exceed a certain limit. If the limit is exceeded a permanent increase in the length occurs. This limit is known as the *elastic limit*.

If we plot the extension of the wire under different loads, we get a curve as shown in fig. 70. The graph is a straight line upto a point A. The load corresponding to this point divided by the cross section of the wire gives the elastic limit. Upto the elastic limit, the extension is proportional to the load, or the stress is proportional to the strain.

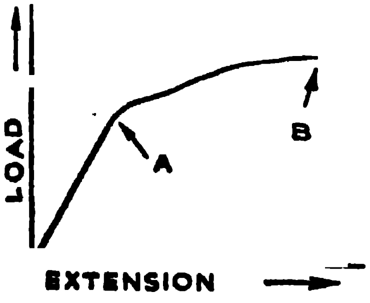


Fig. 70

Hooke's Law. The fundamental law relating to elasticity was discovered by Robert Hooke (1635-1703), an English physicist. It states that **so long as the elastic limit is not exceeded the stress is proportional to the strain.**

$$\therefore \text{Stress} = \text{a constant (say, } K) \times \text{strain.}$$

The constant K has a value characteristic of the material of the elastic body, and is called the *modulus of elasticity*.

Young's Modulus. If e is the elongation produced by a load F when applied to a wire of original length l and cross section $\propto$, then the stress $= F/\propto$ and the strain $= e/l$. The ratio of this stress to the strain is called Young's modulus. Denoting the latter by Y, We have $Y = \frac{F/\propto}{e/l} = \frac{Fl}{\propto e}$

Since strain is a number, Young's Modulus is expressed in the same units as a stress, e.g., in pounds per sq. in., gm. or kg. or dynes per sq. cm. etc.

Some other kinds of strain. The extension of a wire under a load represents a kind of strain known as *tensile strain*. The corresponding stress is called *tensile stress*. If the force is one of compression along a direction instead of extension as above, the corresponding strains and stresses are called *compressive strains* and *compressive stresses*.

When a beam supported at the ends is loaded at the middle, the depression is proportional to the load so long as a certain limiting load is not exceeded. The extension of a loaded spring is proportional to the load producing it unless the load exceeds a limit. Action of the spring balance is dependent on this behaviour of springs.

The twist in a wire is porportional to the couple producing the twist provided the latter does not exceed a certain value.

In all cases the limiting value depends on the nature of the deformation, the material of the body undergoing the deformation and the dimensions of the body.

80. Centre of Gravity

A body may be looked upon as being made up of a large number of particles, each of finite mass. Each particle is attracted towards the centre of the earth by the force of gravity, which is proportional to the mass of the particle. Owing to the large radius of the earth these forces may be taken as parallel.

The weight of a body, therefore, consists of a system of parallel forces acting upon the individual particles which make up the body. The point of application of the resultant of this system of parallel forces is called the *centre of gravity* of the body. The weight of a body may be taken to act at its centre of gravity. It therefore follows that *when a body is supported at its centre of gravity, it will have no tendency to turn under the action of gravity.* The position of the centre of gravity is thus independent of the position of the body.

Centre of gravity of bodies of simple shapes can often be located by inspection. In the case of a thin uniform rod, the centre of gravity (let us henceforth abbreviate it into C.G.) is at the centre of the rod. The turning moment about the centre due to the weight of a particle is balanced by that of another particle similarly placed on the other side of the centre. Hence the total turning moment of the rod about its centre is zero, The centre of the rod is therefore its C.G.

The C.G. of some uniform, homogeneous bodies having simple geometrical shapes are given below:

Body	*Position of C.G.*
Circular lamina	Centre of circle
Annular disc	Centre of annulus
Triangular lamina	Point of intersection of the medians.
Rectangular lamina	Point of intersection of the diagonals
Sphere	Geometrical centre
Spherical shell	Geometrical centre
Cylinder	Midpoint of axis
Cone	On the axis at a distance equal to ¼ of the height above the base.

Experimental determination of C.G. The fact that a body supported through its C.G. has no tendency to turn under the action of gravity, provides a convenient method of finding its C.G. When a body is balanced on a knife edge or, say, a bevelled ruler, its C.G. lies in the vertical plane through the line of balance. If you can find two such lines of balance, the C.G. must be on the vertical line through the point of intersection. If the body is a thin lamina, the point of intersection may be taken as its C.G. It may not, however, be easy to balance a lamina about a knife edge.

Experiment 26. *To find the C.G. of an irregularly shaped lamina.*

Apparatus : Lamina, large pin, plumb-line. (The lamina may have a few small holes around its edge.)

Method : Suspend the lamina by a pin from a point A near its edge so that it can turn about the pin freely. When the lamina is at rest, its C.G. must be vertically below the point A. By means of the plumb line mark the vertical through A on the lamina.

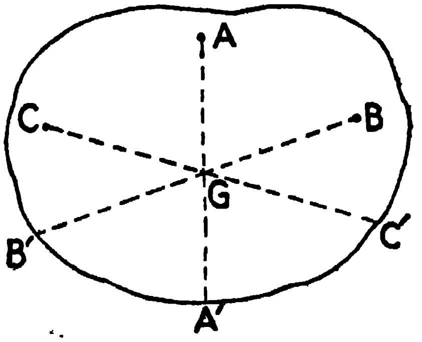


Fig. 71

Change the point of suspension and repeat what you have done in the first case. You will find that the vertical lines through the points of suspension (AA', BB', CC' in Fig. 71) intersect at the same point G. G gives the position of the $C.G.$ of the lamina.

81. Stable, Unstable and Neutral Equilibrium

The behaviour of a body when slightly displaced from its position of rest, depends upon the position of its C.G.

A body is said to be in **stable equilibrium** if, when slightly displaced, it tends to fall back to its original position. Equilibrium is stable when the displacement raises the C.G. A cone standing on its base (Fig. 72), a cube resting on a face, a weight suspended by a string, a ball lying in a spherical cup (Fig. 73), a chair or table standing on its legs etc. are examples of stable equilibrium.

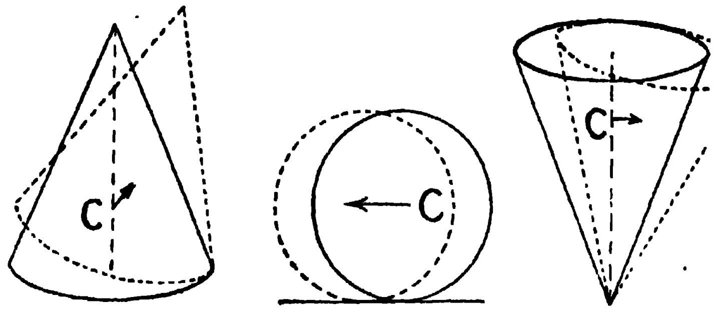


Fig. 72

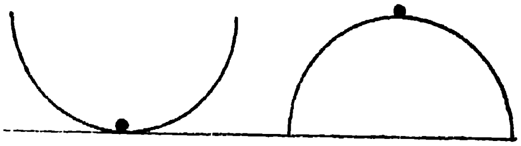


Fig. 73

A body is said to be in **unstable equilibrium** if, when slightly displaced, it tends towards further displacement. Equilibrium is ustable when the displacement lowers the C.G. A chair balanced on two legs, a cube balanced on an edge, a cone balanced on its vertex (Fig. 72), an egg balanced with its long axis vertical, a ball on the top of a sphere (Fig. 73) etc., are examples of unstable equilibrium.

A body is said to be in **neutral equilibrium** if, when slightly displaced, it remains in its new position. Equilibrium is neutral when the displacement neither raises nor lowers the C.G. A sphere resting on a horizontal surface, a pencil, a cylinder or a cone resting on its side etc., are examples (Fig. 72).

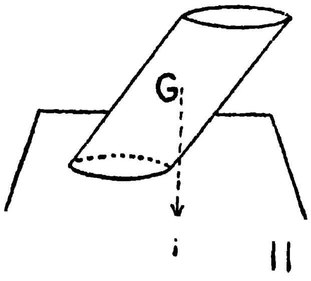


Fig. 74

A body will topple unless the vertical line through its C.G. passes through the base on which it is supported. The cylinder in Fig. 74 cannot be kept standing on an end. Try piling up bricks one above the other, but slightly displaced to one side. Soon the pile will topple. This occurs when the vertical line through the C.G. of the pile moves outside the base offered by the lowest brick. Place a cylinder with its axis vertical on a board, and gradually tilt the board. The cylinder will topple as soon as the vertical line through its C.G. moves out of the base (Fig. 75).

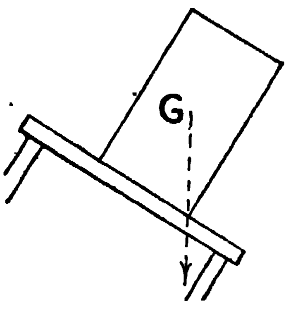


Fig. 75

The stability of a body depends on how far it may be tilted without toppling. High stability is obtained by keeping the C.G. low and making the base large. This is an important point to remember in the design of structures.

Extra passengers should not be allowed on the upper deck of double-decker buses. If they are allowed, the C.G. will rise, and the risk of the bus toppling over while taking a bend will increase. Racing cars are built low for the same reason.

C.G. of a combination of two portions. If a body is made of two portions the C.G. of each of which is known, it is easy to find the C.G. of the body as a whole. In Fig. 76 let W_1 and W_2 be the weights of the two portions making up the body, and C_1 and C_2, their centres of gravity respectively. The centre of gravity of the body as a whole will lie at a point C on the line joining C_1 and C_2 such that

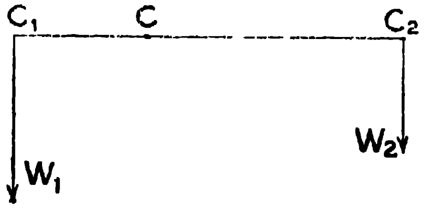


Fig. 76

$$CC_1 \times W_1 = CC_2 \times W_2$$

$$\text{or} \quad \frac{CC_1}{CC_2} = \frac{W_2}{W_1}$$

$$\text{or} \quad \frac{CC_1 + CC_2}{CC_2} = \frac{W_1 + W_2}{W_1}$$

$$\text{or} \quad CC_2 = C_1C_2 \times \frac{W_1}{W_1 + W_2}$$

$$\text{Similarly} \quad CC_1 = C_1C_2 \times \frac{W_2}{W_1 + W_2}$$

If the portions are in the form of uniform laminæ their weights will be proportional to their areas. If, besides, they have simple geometrical shapes the positions of C_1 and C_2 could be obtained by calculation. We thus have data enough to compute the position of C. If the body is a homogeneous solid, weight will be proportional to the volume.

Problem. A uniform lamina is in the form of a square with an equilateral triangle of the same side on top of it (Fig. 77). If the sides are 10 inches each, find where the C.G. of the lamina is.

Solution : C_1, the C.G. of the square, is at the intersection of the diagonals. Hence it is 5 inches below the centre O of the common side AB. C_2, the C.G. of the triangle, is one-third way up the median through O

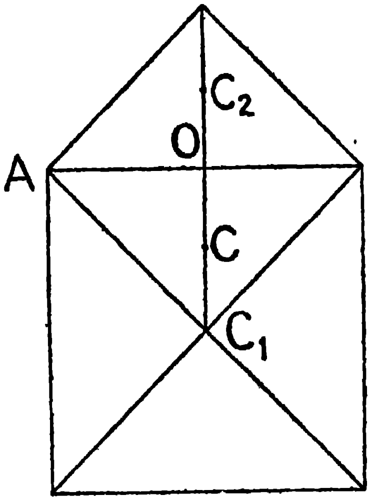


Fig. 77

i.e., $OC_2=\frac{1}{3}\times 10\times\frac{\sqrt{3}}{2}$ inches $=\frac{5}{\sqrt{3}}$ inches. $\therefore C_1C_2=\left(5+\frac{5}{\sqrt{3}}\right)$ in.

Area of square, $A_1=100$ sq. in.; area of triangle, $A_2=\frac{1}{2}\times 10\times 10\times\frac{\sqrt{3}}{2}=25\sqrt{3}$ sq. in.

$$\therefore CC_1=C_1C_2\times\frac{A_2}{A_1+A_2}=\left(5+\frac{5}{\sqrt{3}}\right)\times\frac{25\sqrt{3}}{100+25\sqrt{3}}=2{\cdot}38$$

inches nearly.

Exercises

1. Distinguish between solids, liquids and gases on the basis of 'shape' and 'volume'. What kinds of molecular aggregation do they have ?

2. State the law of gravitation. Show how the acceleration due to gravity is related with the gravitational constant and the mass and radius of the earth.

3. What is the relation between mass and weight ? How does the latter behave at different places on the earth, and above and below its surface ?

4 State the laws of simple pendulum. Describe fully how you would measure g with such a pendulum.

What is the special feature in the motion of the bob of a simple pendulum oscillating with a small arc ?

5. Define stress, strain and Young's modulus. What is Hooke's law ?

Calculate in appropriate units the value of Young's modulus given that a wire of length 1 metre and diameter 1 mm. extends by 1 mm. when stretched by a load of 10 kg.

(*Ans* : $1{\cdot}27 \times 10^6$ kg. per sq. cm.)

6. Define centre of gravity. How does a body behave when supported at the centre of gravity? State the position of the C.G. when the body is (*i*) a uniform triangular lamina, (*ii*) a uniform right circular cone.

7. How would you determine the C.G. experimentally in the case of a lamina?

8. What do you understand by stable, unstable and neutral equilibrium? Give two examples of each.

9. (a) A block of wood measures $9'' \times 5'' \times 3''$. How should it lie so as to have maximum stability?

(b) Why is it dangerous to load passengers on the upper deck of a double-decker bus?

(c) Why are racing cars built low and with a larger distance between wheels than in other cars of similar size?

CHAPTER 10

THERMOMETRY

82. Heat

We derive the sensations of warmth and cold through our sense of touch. When a vessel containing cold water is placed over a fire, it becomes warm, then hot and finally begins to boil. *The external agent which turns a cold body hot is called heat.* When a body is heated we conclude that heat has been added to it.

All bodies contain heat. However cold a body is, it contains some heat. Water in a lake, pond or river is colder than boiling water; ice is colder than water in a lake. If ice is added to a tumblerful of ordinary water, it melts. The water becomes colder than before, but remains warmer than ice. We can explain it by saying that some heat has passed from the water into the ice and melted and warmed it. The tumblerful of ordinary water which was colder than boiling water, undoubtedly contained heat.

Even ice contains heat. Liquid air is much colder than ice. When a vessel containing liquid air is placed over ice, liquid air boils like water over a fire. Heat passes from ice into the liquid air.

The same agent 'heat' is responsible for both the sensations of heat and cold. When heat enters our body from the object touched, we feel the sensation of warmth. When heat leaves our body, we feel the sensation of cold.

83. Temperature

We have said that all bodies contain heat. It should be remembered that our sensation of warmth or cold does not depend on the total amount of heat that a body contains. When a quantity of cold water is poured into a vessel containing hot water, the total amount of heat in the water of the vessel is larger than before ; but the water appears colder than it was.

If the sensation of warmth is not to depend on the total quantity of heat in a body, it must depend on some other factor. This factor is called *temperature*. In trying to define temperature we observe that when two bodies are kept in contact, heat may pass from one body into the other. One of the bodies gets colder and the other warmer. The former loses heat to the latter. This behaviour of bodies helps us in defining temperature.

Temperature is that thermal condition of a body which determines its ability to transfer heat to other bodies. Heat flows from the body at the higher temperature to the body at the lower temperature. When the temperature is the same for two bodies heat ceases to flow from one to the other.

If we take a cupful of water out of a bath tub, the water in the cup has the same temperature as that in the tub. But as the tub may contain a thousand cupfuls of water, the total heat in the tub is much greater than that in the cup.

84. Difference between Heat and Temperature

The points of difference between heat and temperature should be clearly realized.

(i) Addition of heat to a body causes its temperature to rise. Heat is the 'cause', and rise of temperature the 'effect'.

(ii) Temperature of a body provides no measure of the heat that the body contains. The temperatures of two bodies may be equal, yet they may contain widely different quantities of heat. A bucketful of water and a cupful taken out of it will serve as an example.

The quantity of heat in a body depends on its temperature and mass, as well as on the substance of which it is made.

(iii) The rise in temperature of two bodies may be different even when the same heat is supplied to both. A cup of boiling water added to a bucket of water raises its temperature but little (makes it slightly warmer). But the same cupful warms a tumbler of water from the bucket much more.

(iv) The flow of heat from one body to another is determined by their temperatures and not by the amounts of heat that they contain. Heat flows from the body at the higher temperature to that at the lower temperature.

The relation between heat and temperature is very similar to that between water and its level. When water is put in a vessel, the level of water in it rises. When heat is put into a body, its temperature rises. Temperature is, therefore, a kind of 'heat level'.

When two vessels in which water stands at different levels are connected together, water flows from the vessel which has the higher level to that at the lower level until the level is the same in both. When two bodies at different temperatures are brought into contact heat flows from the body at the higher temperature to that at the lower, until the temperature is equalized. The direction of flow is not determined by the total quantity of water in the vessel or the total amount of heat in the body.

85. Effects of Heat

The effects of heat are too many to enumerate. Almost all physical properties are affected more or less by heat. It can also bring about chemical changes. Most chemical actions take place faster at higher temperatures. This also applies to life processes ; but life cannot continue at high temperatures. In this book we shall however concentrate on the following effects of heat, viz., (i) rise in temperature, (ii) change of size, and (iii) change of state from solid to liquid and liquid to vapour. Heat raises the temperature ; the other changes are direct effects of the rise of temperature. Incandescence is such an effect. Bodies when sufficiently hot emit light. A candle flame contains incandescent carbon particles. In an electric bulb the filament has been brought to incandescence by electric current.

86. Thermometers

A thermometer is a device for measuring temperature. Our sense of touch does not provide an accurate means for measuring temperatures. It may even give us a wrong idea. If one hand is dipped in hot water, and the other in cold water, and then both hands in tepid water, the hand that was dipped in hot water feels cold, while the other feels warmer. When the difference of temperature is small we may fail to realise the difference by our sense of touch. Besides, our sense of warmth or cold is also determined by the rate at which heat flows into or out of our body. On a hot day a piece of metal feels warmer than a piece of wood ; but on a cold day the sensations are reversed, though on any day the temperature of the metal was the same as that of the wood.

It thus becomes necessary to devise a method of measuring temperature. We know that the volume of a substance increases with rise in temperature. The increase is much larger for liquids than for solids. The expansion of a liquid with rise in temperature may be utilized for measuring temperatures. For various reasons (see § 90) mercury is preferred to other liquids for temperature measurement.

The most common form of thermometer is the mercury-in-glass thermometer. In it the excess expansion of mercury over that of the glass bulb in which it is contained is utilized for the measurement of temperature. The construction of a mercury-in-glass thermometer is described below.

87. Construction of a Mercury Thermometer

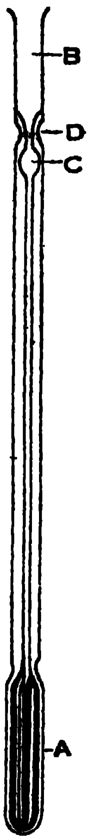


Fig. 78

A mercury-in-glass thermometer can be constructed from a long piece of capillary glass tube which has a bulb *A* and a funnel *B* (Fig. 78) sealed at its two ends. A constriction is made in the tube at *D* below the funnel, while just below it there is a small bulb *C*. The interior of the tube must be free from dust, grit and grease.

Mercury is placed in the funnel and the bulb *A* gently heated. Air in *A* expands and partly escapes through the mercury. As *A* cools a little mercury is drawn into the bulb. If this mercury is now boiled mercury vapour drives the air entirely out of the bulb and the tube. On cooling the vapour condenses ; more mercury from the funnel is drawn into *A* and completely fills it.

The bulb is then maintained at a temperature a little above the highest to which the thermometer will ever be used and finally sealed off at *D*. The small bulb *C* serves as a protective device. If ever the thermometer is heated to a temperature higher than that for which it is designed mercury will collect in *C* and prevent the thermometer from breaking.

The fixed points of a thermometer. The thermometer so constructed is not yet ready for use. A scale for measuring temperatures has to be fitted to it. For this purpose we must determine the positions (within the tube) which the mercury meniscus occupies at two known temperatures. These two temperatures are the temperature of melting ice and that of boiling water. The former is called the *ice point* or the *lower fixed point* of the thermometer, and the latter, the *steam point* or the *upper fixed point* of the thermometer. These two fixed points have the advantage that they are readily reproducible anywhere in the world.

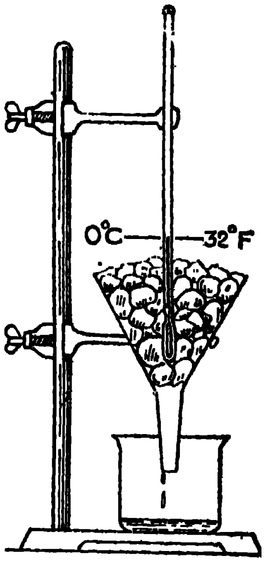


Fig. 79

Experiment 27. *To determine the lower fixed point of a thermometer.*

Apparatus. An ungraduated mercury thermometer, some crushed ice, a glass funnel, a beaker and a stand.

Place the thermometer in a funnel as in (Fig. 79). The bulb and the lower part of the stem must be completely surrounded by finely crushed melting ice. From time to time observe the position of the mercury meniscus in the tube. When the thermometer acquires the temperature of melting ice, the level of mercury in the tube maintains a steady position. Note this position. This is the lower fixed point.

Experiment 28. *To determine the upper fixed point of a thermometer.*

Apparatus. An ungraduated thermometer, a hypsometer and a bunsen flame.

The thermometer is kept in a steam bath. This may be conveniently done with the help of a hypsometer (Fig. 80), which consists of a cylindrical vessel A in which water is boiled. The steam passes up the cylinder BB then down the region between the two cylinders BB and CC and out at D. The thermometer is inserted through a cork in the cover and so arranged that the final level of mercury is just visible.

When the thermometer has acquired the temperature of steam the level of mercury acquires a steady position, which is marked. This is the upper fixed point of the thermometer.

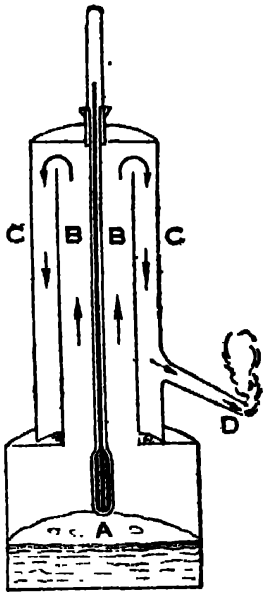


Fig. 80

88. Scales of Temperature

After the fixed points of a thermometer have been determined, the distance between them is divided into a number of equal parts so as to form a scale for measuring temperature. There are two scales in common use.

(i) **The Centrigrade Scale.** On this scale the temperature of melting ice is taken as zero degree centigrade (written $0°C$) and that of steam boiling under normal atmospheric pressure as $100°C$. When the space between the lower and the upper fixed points of a thermometer is divided into 100 equal parts, each part corresponds to a difference of temperature of $1°C$. This scale of temperature is used in scientific measurements all over the world. Many countries, of which India has recently become one, use this scale for domestic, industrial and commercial purposes also.

(ii) **The Fahrenheit Scale.** On this scale the ice point is taken as 32°*F* and the steam point as 212°*F*. The distance between the two fixed points is divided into 180 equal parts. Each part corresponds to a temperature difference of 1°*F*. The scale is used for domestic, industrial and commercial purposes in the U.S.A., the U.K., and other countries of the commonwealth except India. The doctors' thermometer is graduated on this scale.

Thermometers graduated on the centigrade and Fahrenheit scales are represented in fig. 81.

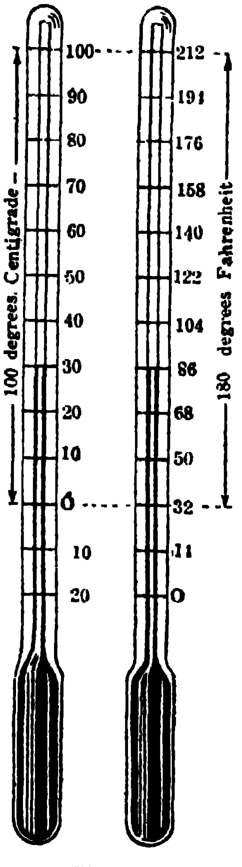


Fig. 81

89. Relation between the two Scales of Temperature

We find from the definition of the scales that—

a temperature difference of 100°*C*
= a difference of 180°*F*.

This helps us in converting a reading from one scale to the other. Suppose that the reading of the same temperature on two thermometers are *c* on the centigrade scale and *f* on the Fahrenheit scale. A temperature which reads *f*° on the Fahrenheit scale is $(f-32)°F$ away from ice point.

Since a difference of $180°F$ = a difference of $100°C$,

„ „ „ $(f-32)°F =$ „ $\frac{100}{180}(f-32)°C$

$$=\frac{5}{9}(f-32)°C$$

That is, the temperature of $f°F$ is $\frac{5}{9}(f-32)°C$ from the ice point. But this temperature is also $c°C$ from the ice point.

$$\therefore \quad c=\frac{5}{9}(f-32) \qquad (89{\cdot}1)$$

$$\text{and} \quad f=\frac{9}{5}c+32 \qquad (89{\cdot}2)$$

90. Mercury as a Thermometric Substance

A **thermometric substance** is one, the change in some property of which is utilized for measuring temperature. In mercury-in-glass thermometers mercury is the thermometric substance. Temperature is measured by the change in the length of a column of mercury due to the change in volume which the mercury in the bulb undergoes with change of temperature.

The advantages of mercury as a thermometric substance are the following ;

(i) Mercury freezes at $-39°C$ and boils at $357°C$. It can thus be used over a fairly wide range.

(ii) Mercury has a fairly large expansion for a small change in temperature. Moreover, the expansion is fairly constant at different temperatures. *Equal scale divisions therefore represent practically equal changes of temperature.*

(iii) Mercury is a good conductor of heat. When put in contact with a body it quickly acquires the temperature of the latter.

(iv) It is opaque and can be easily seen through glass.

(v) Compared with any other liquid of the *same mass* it takes much less heat from the body of which the temperature is to be measured. But its high density neutralizes this advantage, as a given volume of it has a much larger mass than any other liquid.

(vi) It does not stick to glass.

(vii) It can easily be obtained pure.

91. Maximum and Minimum Thermometers

For meteorological purposes as well as in some industries it is necessary to know the maximum and the minimum temperatures reached at any place in the course of the day. Self-recording thermometers are used for this purpose. These are provided with simple devices which enable one to read off from the thermometer the maximum or the minimum temperature it reached last. Some of these thermometers are described below:

(i) **Rutherford's Maximum Thermometer.** It is a mercury thermometer from which one can read off the maximum temperature it reached since it was set last.

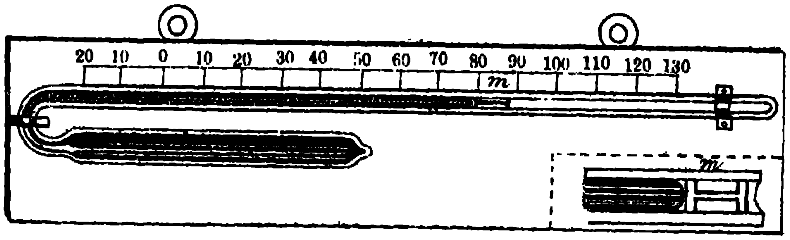


Fig. 82

The thermometer tube is horizontal and has a light steel pointer *m* outside the mercury thread (Fig. 82 and inset). The pointer can be moved with the help of a magnet or by tilting the frame which carries the thermometer. It is placed with one of its ends in contact with the convex surface of the mercury thread. As mercury expands on rise of temperature it pushes the pointer ahead of it. When mercury contracts the column shrinks, leaving the pointer at the place to which it had been pushed. The position of the end of the pointer facing the mercury thread gives the maximum temperature reached.

(ii) **Rutherford's Minimum Thermometer.** To get the minimum temperature Rutherford devised an *alcohol thermometer* which has a light glass pointer (*a*; fig. 83) placed within the liquid.

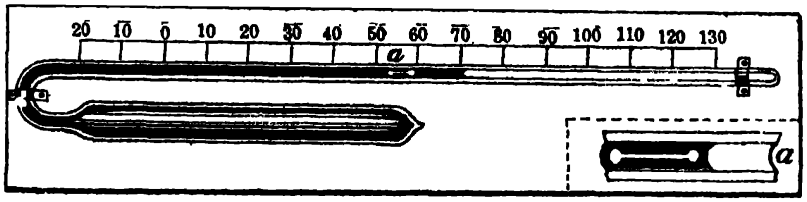


Fig. 83

In use, the thermometer tube is kept horizontal with one end of the pointer in contact with the concave end of the alcohol thread. When the column shrinks with fall of temperature, the pointer is carried by the liquid as the surface of a liquid behaves like an elastic skin. But when temperature rises alcohol flows past the pointer, which is left in the position in which it was. The pointer thus gives the minimum temperature reached.

(iii) **Clinical or Doctors' Thermometer.** It is a maximum thermometer and is used for measuring the temperature of the human body. There is a constriction above the bulb (Fig. 84) which contains mercury. When put in close contact with the body the temperature of the bulb increases. As the mercury expands it forces itself through the constriction and rises up the stem. When removed from contact with the body the temperature of the bulb falls and the mercury shrinks into the bulb. But the constriction being very narrow the mercury thread breaks there. While the mercury below the constriction shrinks into the bulb, that above it remains within the stem and indicates the maximum temperature reached. When a fresh reading has to be taken this thread must be forced into the bulb by jerking. If the temperature of air is higher than that of the human body the mercury in the thermometer continues to rise after it has been removed from contact with the body.

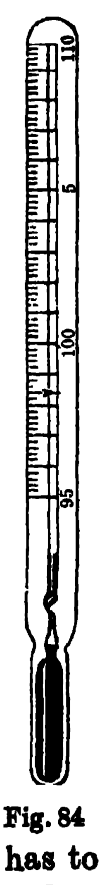


Fig. 84

The scale is graduated in degrees Fahrenheit and has the range 95°F to 110°F. Human life cannot last long beyond this range of temperature. Hence a bigger range is unnecessary, A '½ minute' thermometer means that its bulb should be placed inside the mouth for at least ½ minute to reach the temperature.

Exercises

1. Distinguish between heat and temperature.

2. Describe the construction of a mercury-in-glass thermometer.

What purpose is served by the small bulb at the upper end of the capillary of such a thermometer ?

3. What are the fixed points of a thermometer ? Describe the centigrade and Fahrenheit scales of temperature.

How can a reading in °*F* be converted into °*C* ? Normal temperature of the human body is 98·4 °*F*. How much is it in °*C* ?

4. What are the advantages of using mercury in thermometers ?

5. How would you determine the maximum or the minimum temperature reached at a place in a certain period of time ? Draw a sketch indicating the position of the pointer relative to the liquid thread in the thermometer.

6. Describe the action of a clinical thermometer and say how it can be considered a maximum thermometer.

7. Convert (i) 50°*C*, 90°*C* and 125°*C* into degrees Fahrenheit and (ii) 158°*F*, 203°*F* and 266°*F* into °*C*.
(*Ans* : (i) 122, 194, 257 ; (ii) 70, 95, 130.)

8. Draw a graph to convert °*C* into °*F*, and find from it the value of 45°*C* in °*F*, and of 59°*F* in °*C*. (*Ans* : 113°*F*, 15°*C*.)

9. Express a rise of temperature of 55°*C* in °*F*, and of 117°*F* in °*C*. (*Ans* : 99 ; 65)

10. At what temperature are the readings on the centigrade and Fahrenheit scales the same ? (*Ans* : −40°*C* or °*F*).

11. The ice point on a thermometer reads 20°, and the steam point, 170°. What is the value in °*C* of a temperature which reads 80° on the thermometer ? (*Ans* : 40).

12. Define the absolute zero of temperature.. its value on the centigrade scale and convert the Farenheit scale.

13. Mention four thermometric substances with th

EXPANSION OF SOLIDS

92. Some Demonstration Experiments

When a solid is heated it expands ; when it is cooled it contracts. The expansion or contraction is so small that it cannot be seen with the unaided eye. It therefore requires carefully designed experiments to demonstrate the effect.

Solids expand on heating. (i) Expansion of a solid on heating can be easily demonstrated with the arrangement shown in fig. 85.

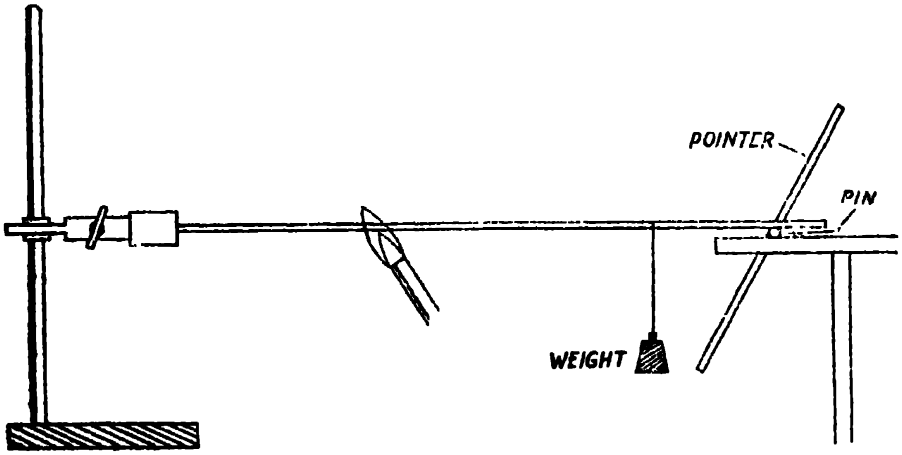


Fig. 85

One end of a thin metal rod is clamped to a stand. The free end is placed on a pin which carries a light pointer. A weight is hung from near the free end of the rod so that the rod presses on the pin and keeps the pointer in position. The rod is heated by running a flame along it. It will be seen that the pointer is deflected from its position. When the rod is allowed to cool the pointer attains its original position. The explanation lies in the expansion of the rod due to heating and subsequent contraction on cooling.

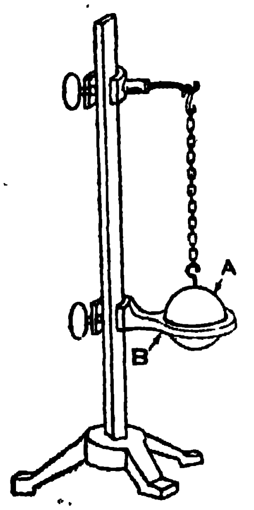


Fig. 86

(ii) *Ball and ring experiment*: Fig. 86 shows a ball *A* which, when cold, just passes through the ring *B*, but does not do so when heated.

(iii) *Expansion of wire when heated electrically.* A nickel wire about ·5mm in diameter is hung from a hook and stretched vertically by a small weight. An electric current of several amperes is passed through the wire so that it becomes red hot. As the wire is heated the weight descends rapidly. It rises when the current is stopped. This provides a vivid demonstration of the expansion of a wire on heating.

Different solids expand differently. Different solids expand by different amounts under identical conditions. This may be shown in the following way.

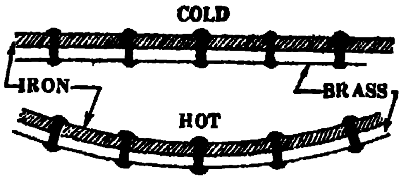


Fig. 87.

Two metal strips of exactly equal length are riveted together (Fig. 87). One may be of copper or brass and the other of iron. On heating, this compound bar bends with the brass or the copper on the outside. On cooling it straightens again. The bending is due to the fact that one strip becomes longer than the other on heating. Brass or copper expands more than iron under the same conditions.

Force of expansion or contraction is enormous. The breaking bar experiment (Fig. 88) demonstrates that an enormous force is brought into play when a solid contracts. A stout iron rod fits across two slots in a heavy iron stand. The rod has a hole through one end and a screw at the other. A cast iron pin (about $\frac{1}{4}$ inch in diameter) is passed through the hole. The rod is then heated and clamped tightly while hot by means of the screw. When the rod is allowed to cool it is found that the cast iron pin snaps owing to the enormous force of contraction. A slight modification of the rod (using another hole) enables us similarly to demonstrate the force of expansion.

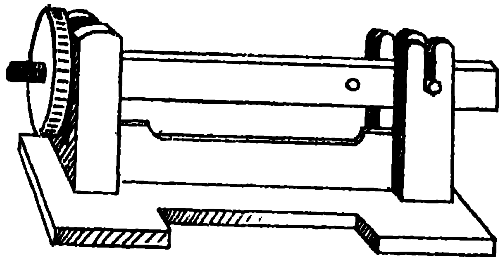


Fig. 88

93. Coefficient of Linear Expansion

When a solid is heated it expands in all directions ; but very often we are concerned with its expansion only in one direction, viz., length. We speak of this as the *linear expansion* of the solid.

Let l_1 be the length of a bar at a temperature t_1 and l_2, its length at temperature t_2. Experiment shows that the increase in length, $l_2 - l_1$, is approximately proportional both to l_1 and to the rise in temperature $t_2 - t_1$. We may, therefore, write

$$l_2 - l_1 = \alpha l_1 (t_2 - t_1)$$

where α is a small constant whose value depends on the material. It is called the coefficient of linear expansion of the material. It should properly be called the mean coefficient of linear expansion between the temperatures t_1 and t_2. Writing

$$\propto=\frac{l_2-l_1}{l_1(t_2-t_1)} \qquad (93\cdot1)$$

we may define $\propto$ as follows: *The coefficient of linear expansion is the change in length per unit length per degree rise of temperature.*

Coefficient of linear expansion

$$=\frac{\text{Increase in length}}{\text{Original length}\times\text{rise in temperature}}$$

It follows that if l_1 is the length at a temperature t_1 of a piece of material of coefficient of linear expansion $\propto$, its length l_2 at temperature t_2 will be given by

$$l_2=l_1\left\{1+\propto(t_2-t_1)\right\} \qquad (93\cdot2)$$

It is easy to see that $\propto$ is independent of the unit of length, but is dependent on the temperature scale. In equation 93·1 $(l_2-l_1)/l_1$ represents the change in length per unit length and is a pure number. In mentioning the value of $\propto$ we must say if the change in length was caused by 1°*C* or 1°*F* rise of temperature. $\propto$, therefore, depends on the scale of temperature adopted.

The following example will make it clear. Let the lengths of a bar at the temperatures of melting ice and of boiling water be l_1 and l_2 respectively. Adopting the centigrade scale and denoting by $\propto_c$ the appropriate coefficient of linear expansion we have

$$\propto_c=\frac{l_2-l_1}{l_1\times100}$$

On the Fahrenheit scale we have the corresponding coefficient given by

$$\propto_F=\frac{l_2-l_1}{l_1\times180}$$

$$\therefore\quad \propto_F=\frac{5}{9}\propto_c$$

The meaning of a statement like "the coefficient of linear expansion of brass is 19×10^{-6} per °*C*" should be clearly understood. It means that for 1°*C* rise of temperature a rod of brass expands by 19×10^{-6} part of its original length. In other words

for 1°*C* rise of temperature 1 cm long brass rod increases by 19×10^{-6} cm
" " " " 1 metre " " 19×10^{-6} metre
" " " " 1 yard " " 19×10^{-6} yard
" " " " l units " " $19\times10^{-6}\times l$ units.

94. Determination of the Coefficient of Linear Expansion

In all solids the expansion is so small that we must either magnify the change in length before measuring it or use some delicate measuring instrument. There are many forms of apparatus, of which we shall describe only one which uses a micrometer screw, and is known as *Pullinger's apparatus* (Fig. 89).

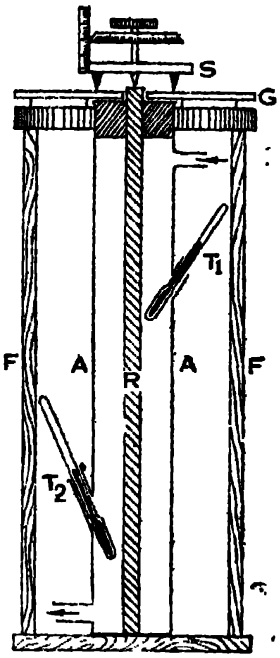


Fig. 89

The experimental rod R stands vertically in a tube A mounted on the frame F.A has two thermometers T_1, T_2 inserted in it to measure the temperature of R. It has also side tubes for inlet and outlet of steam and is surrounded by felt or some other non-conductor of heat. The top of R passes through a cork in A and through a hole in a glass plate G which rests horizontally on the frame. A spherometer rests on G. After the initial length l of the rod has been measured it is placed in position and the central leg of the spherometer brought into contact with the top of R (This adjustment may be accurately made by connecting one end of an electric circuit containing an electric bell with a leg of the spherometer and the other end with R. When the contact between the spherometer and the rod R is established the bell starts ringing). The spherometer reading is taken. The central leg may now be raised without removing the spherometer. The initial temperature t_1 of the rod is read off from T_1 and T_2.

Steam is then allowed to pass through A. Temperature of R as indicated by T_1 and T_2 rises and finally becomes steady. When it has been steady for some time the central leg of the spherometer is again brought into contact with R and its reading taken. The difference in the two readings of the spherometer gives the expansion x of the rod. The final temperature t_2 of the rod is noted from T_1 and T_2. We have all the data to find the coefficient of expansion α.

$$\alpha = \frac{x}{l(t_2-t_1)}$$

95. Values of α for some Substances

The following table gives the values of the coefficient of linear expansion of some common substances on the centigrade scale.

They are mean values between room temperature and steam point.

Substance	α per °C	Substance	α per °C
Aluminium	$25\cdot5\times10^{-6}$	Brass	$18\cdot9\times10^{-6}$
Copper	$16\cdot7$,,	Bronze	$17\cdot7$,,
Iron (cast)	$10\cdot2$,,	Glass (crown)	$8\cdot5$ to $9\cdot7$,,
,, (steel)	$10\cdot5$ to $11\cdot6$,,	,, (flint)	$7\cdot8$,,
Platinum	$8\cdot9$,,	,, (pyrex)	3 ,,
Silver	$18\cdot8$,,	Invar (64Fe, 36Ni)	$0\cdot9$,,
Zinc	$26\cdot3$,,	Quartz (fused)	$0\cdot5$,,

The International Prototype Metre, made of an alloy of 90% Pt and 10% Ir, has a coefficient of expansion of $8\cdot7\times10^{-6}$ per °C. The Imperial Standard Yard, made of 32 parts of copper, 2 parts of zinc and 5 parts of tin, has a coefficient of $17\cdot7\times10^{-6}$ per °C.

96. Coefficient of Surface and Volume Expansion

When the length of a body changes with temperature, the area of its surface as well as its volume undergoes changes.

The *coefficient of superficial expansion* is the change in area per unit area per degree rise of temperature. If S_1 and S_2 are the areas of any portion of the surface of a body at temperatures t_1 and t_2 respectively, then the coefficient of superficial expansion

$$\beta=\frac{S_2-S_1}{S_1\,(t_2-t_1)}$$

The *coefficient of cubical expansion* is the change in volume per unit volume per degree rise of temperature. If V_1 and V_2 are the volumes of a body at temperatures t_1 and t_2 respectively, then the coefficient of cubical (or volume) expansion

$$\gamma=\frac{V_2-V_1}{V_1\,(t_2-t_1)}$$

Like the coefficient of linear expansion, these coefficients also depend on the temperature scale only and are expressed in per °C or per °F as unit.

Relation between α, β and γ :

Consider a square of sides l cm. Let it be heated through 1°C. Each side becomes $l(1+\alpha)$ cm and the area $l^2(1+\beta)$ sq. cm $\therefore\ l^2(1+\beta)=l^2(1+\alpha)^2$ or $1+\beta=1+2\alpha+\alpha^2$ or $\beta=2\alpha$ (α^2 being neglected).

If a cube of sides l cm. be heated through 1°C, the sides become $l\,(1+\alpha)$ cm. and the volume is changed to $l^3\,(1+\gamma)$ c.c.
$\therefore\ l^3(1+\gamma)=l^3(1+\alpha)^3$ or, $1+\gamma=1+3\alpha+3\alpha^2+\alpha^3$ or, $\gamma=3\alpha$ (α^2, α^3 being neglected)
[Since α is very small, α^2, α^3 are smaller still. For example, if α is ·00001, then $\alpha^2=$·0000000001 and $\alpha^3=$·000000000000001.

We are, therefore, justified in neglecting α^2 and α^3 compared with α.]

97. Practical Effects of Expansion of Solids

A. Expansion of solids serving a useful purpose:

(i) Metal tyres of cart wheels are fitted on when hot. On cooling they contract and attach rigidly to the wheels. Locomotive wheels made of iron are fitted with steel tyres in the same way.

(ii) Steel plates of boilers, ships or bridges are held together by rivets. Rivets are fixed while hot. On cooling they contract and make a very firm joint (Fig. 90).

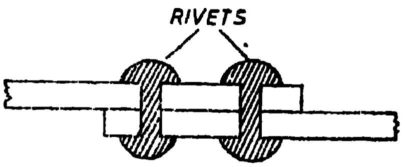


Fig. 90

(iii) Walls which have bulged outwards may be drawn in by passing iron bars through them across the building. To the extremities of the bars are attached iron plates screwed upto the walls with nuts. The bars are heated and the nuts and plates screwed up tightly against the wall. When the bars contract they draw the walls with them.

(iv) Glass stoppers often stick to bottles too tightly to be opened by ordinary means. Careful heating of the mouth of the bottle enlarges it, when the stopper comes out easily. If a metal cap sticks, the cap itself is heated.

B. Disadvantageous expansions and their remedy :

Expansion of solids more often cause inconvenience and calls for suitable remedies.

(i) The rails on a railway expand on a hot day and contract at night. The extremes of temperature may be fairly large. To allow for the expansion and contraction of the rails a small gap is always left between the ends of adjacent rails. The arrangement is shown in fig. 91. The rails are joined by fish plates bolted to the rails. Bolt holes are oval in shape and allow the rails to move in the direction of their length.

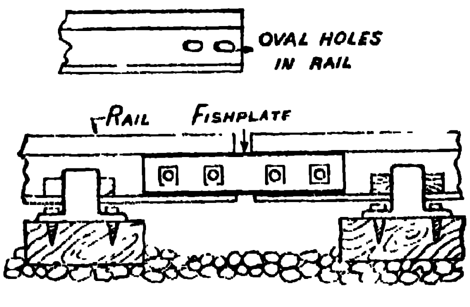


Fig. 91

In tramways however it is the common practice to weld rails together in order that the resistance to the flow of electric current through the rails to the earth may be the least. The rails are imbedded in the road and surrounded by granite blocks and concrete. This gives the rails much protection against temperature changes. The rails of a railway are fully exposed.

(ii) The ends of a big steel bridge are not rigidly built into the brick work on which they rest. They are supported on rollers which allow them to expand or contract without damaging the masonry (Fig. 92).

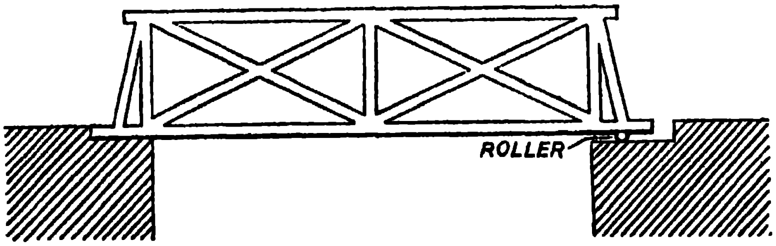


Fig. 92

(iii) When hot water is poured in a glass vessel with thick walls, the vessel is likely to crack. The inner wall tends to expand; but as glass is a bad conductor of heat the outer wall remains cooler and does not expand. This sets up large stresses in the glass, which cracks. The risk is less when the wall is thin or when the glass has a small coefficient of expansion. The linear coefficient for ordinary glass is about 9×10^{-6} per °C. Pyrex glass has a coefficient of 3×10^{-6} per °C. It can be heated in a flame without cracking. Fused silica (quartz) has such a small coefficient that a crucible made of it can be heated red hot in a furnace and chilled in water without cracking.

Unequal expansion of glass and many metals makes it impossible to seal a metal wire such as copper or iron into a glass bulb. Platinum and some alloys have very nearly the same expansion as glass, and are used for sealing. In modern electric lamps the leads which pass through the glass stem are made of an alloy of nickel and iron coated with copper. This alloy expands equally with the glass.

(iv) To allow for the expansion and contraction of steam pipes they are provided with expansion bends or loops as shown in fig. 93. When the pipe expands or contracts the shape of the loop changes slightly. The pipes carrying oil through deserts are also fitted with such bends and loops to allow for their expansion and contraction. They may also be laid in zig-zag fashion loosely on supports.

(v) A telegraph, telephone or a power line is fixed between poles in such a way as to allow for its expansion and contraction.

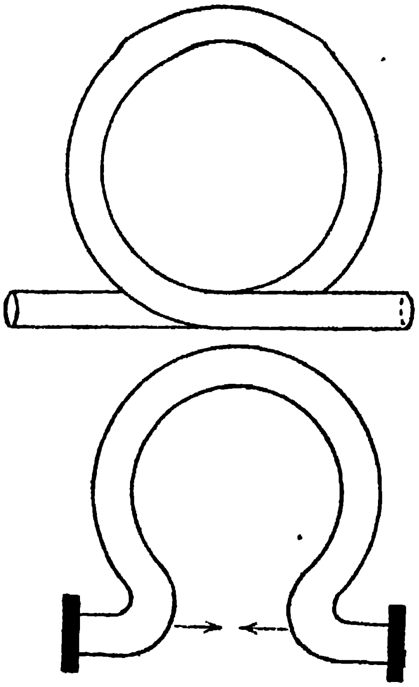


Fig. 93

(vi) The time period of a pendulum depends on its length. Since the length increases in summer, the time period also increases and the clock goes 'slow'. The reverse takes place in winter. Arrangement must be made in order that a clock or a watch may keep correct time in all seasons. Such clocks or watches are called 'compensated'. The effective length of a pendulum is the distance between the point of suspension and the centre of gravity of the suspended system. If this length remains constant the pendulum keeps correct time. The pendulum bob is mounted on rods of two different materials in such a way that the downward expansion of one rod is equal to upward expansion of the other. This is diagrammatically shown in Fig. 94. Let l and l' be the lengths of the rod and α and α' their coefficients of linear expansion. Their expansions for $t°$ rise of temperature are $l\alpha t$ downwards and $l'\alpha' t$ upwards. The distance of the centre of gravity of the pendulum from the point of suspension will remain unchanged if $l\alpha t = l'\alpha' t$

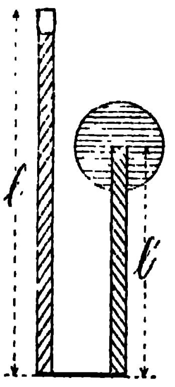


Fig. 94

$$\text{or,}\quad l/l' = \alpha'/\alpha$$

The coefficients of iron and zinc are ·00001 and ·000028 per °C respectively. Hence zinc rods used for compensating iron rods will have less than half the length of the latter. The clock 'Big Ben' on the Parliament House in London uses zinc for compensation. Its pendulum is as shown in the Fig. 95. Smaller clocks may have the pendulum rod made of 'invar', an alloy of 64% iron and 36% nickel. It has an expansion coefficient of $\cdot 9\times10^{-6}$ per °C (about 1/13 that of iron). The very small expansion of the invar rod is compensated by a brass tube (Fig. 96).

Compensated balance wheels of watches. In watches the movement of the hands is controlled by balance wheels which execute rotational oscillations under the influence of a steel hair-spring. The time of oscillation depends on the effective distance of the rim of the wheel from the axis of rotation. The rim is made of two metals, the outer being more expansible. This double rim is made in two or three parts (Fig. 97) each supported by an arm of the wheel. The rim also carries small screw weights. When the temperature rises the arm expands, but the outer rim of more expansible metal curls in and brings the weights near to the axis, thus keeping the effective radius of the ring constant. With change of temperature the stiffness of the hair-spring changes and alters the time-period. The rim is arranged to compensate for this effect also.

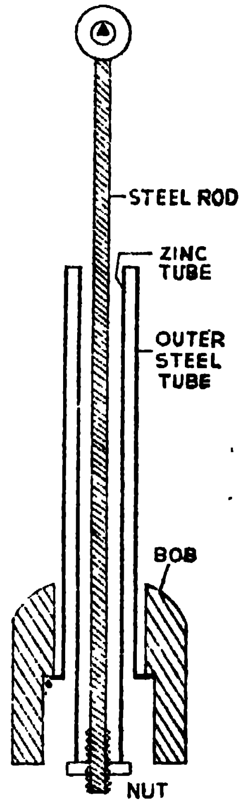


Fig. 95

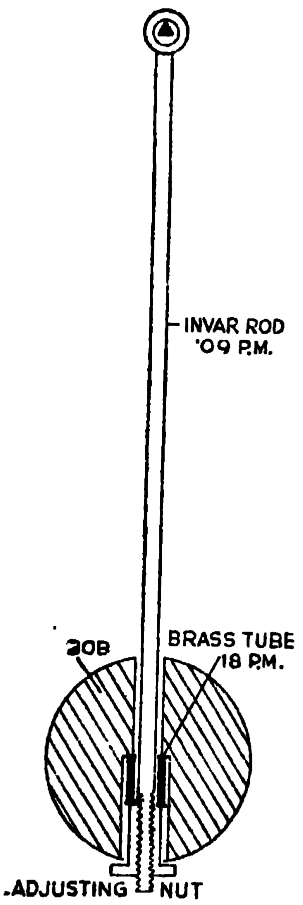


Fig. 96

Strips made of two different metals riveted together curl when temperature changes. Such bimetallic strips may be used as thermometers. The method of use is illustrated in fig. 98.

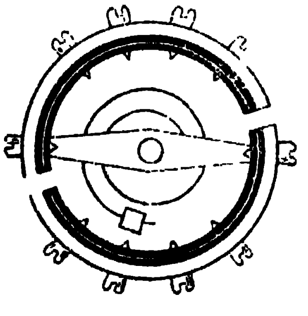


Fig. 97

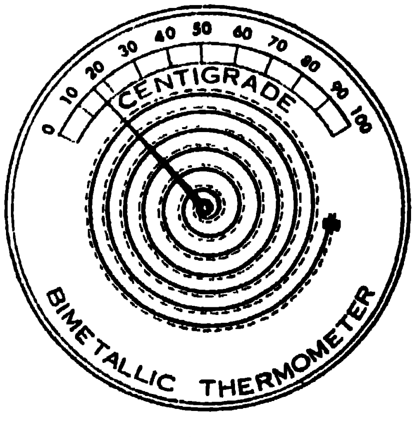


Fig. 98

The action of one type of fire alarms depends on the bending of bimetallic strips on heating. When the heat of the fire causes the strip to curl, it closes an electric circuit which rings the alarm. Such fire-alarms may be installed in ware-houses, in holds of ships or in other places where a fire may break out without being quickly discovered.

Exercises

1. Define coefficient of linear expansion. What is the unit in which it has to be expressed ? How does its value change if the temperature scale is changed from centigrade to Fahrenheit ?

2. Show how the coefficient of volume expansion is related to that of linear expansion.

3. Give three examples where expansion on heating is made to serve a useful purpose.

4. Give three examples where expansion on heating causes inconvenience, and say how it is remedied.

How can pendulums be constructed to have an invariable length ?

5. An aluminium bar is 8 ft. long at $35^\circ C$. To what temperature must it be heated to expand by ·04 in. α for aluminium = ·000026 per $^\circ C$ (*Ans* : $51^\circ C$).

6. Find the temperature to which a zinc rod 50 cm. long at $59^\circ F$ must be heated in order that its length may be 50·05 cm. α for zinc = ·000025 per $^\circ C$ (*Ans* : $131^\circ F$)

7. α for steel is ·000012 per $^\circ C$. What is its value in $^\circ F$? If the length of a steel rail laid at $80^\circ F$ is 5 yds., what interval must be left between the rails to allow for a rise in temperature to $112^\circ F$. (*Ans* : 0·0384 in.)

8. The span of a steel bridge is 1250 ft. Assuming α = ·000012 per $^\circ C$, calculate the change in length of a span when the temperature changes from $30^\circ C$ to $40^\circ C$ (*Ans* : ·15 ft.)

9. A 15 ft. long girder increases by ·0324 in. when temperature changes from $40^\circ C$ to $55^\circ C$. Find the coefficient of linear expansion of the material of the girder. (*Ans* : .000015 per °C).

10. The diameter of a sphere at $35^\circ C$ is 1 in. Find the temperature to which it should be raised so that it will just fail to pass through a circular hole of diameter 1·001 in. Coefficient of cubical expansion of the material of the sphere is ·000075 per $^\circ C$. (*Ans* : $75^\circ C$).

11. A brass metre scale is correct at 30°*C*. The length of a table measured by it at 55°*C* is 40 cm. What is the true length? α for brass is ·00002 per °*C* (*Ans*: 40·02 cm.).

12. A wheel has a diameter of 75 cm. A steel tyre of inner diameter 74·8 cm. is to be fitted on it. Calculate the range of temperature in degrees centigrade through which the tyre must be heated so that it may be mounted on the wheel. α for steel is ·000012 per °*C*. (*Ans* : 222·8°*C*)

13. Two bars of brass and steel, standing side by side, have one end of each rigidly attached to the other. The other ends are free to expand. The steel bar is 1 metre long. What should be the length of the brass bar so that the distance between the free ends of the bars remains the same at all temperatures? α for brass is ·00002 per °*C*, and α for steel is ·000012 per °*C*. (*Ans* : 60 cm.).

CHAPTER 12

EXPANSION OF LIQUIDS

98. Real and Apparent Expansion of a Liquid

A liquid has no shape of its own; it readily takes the shape of the container. We cannot therefore speak of the linear or superficial expansion of a liquid. Liquids have volume expansion only.

When a liquid is heated the container is also heated along with it and expands. The expansion of the vessel masks the expansion of the liquid and makes the latter appear smaller than its real value. In other words, the apparent expansion of a liquid (i.e., the expansion of a liquid that we observe) is less than its real expansion.

To show the expansion of a liquid and that of the containing vessel: The expansion of the container may be demonstrated as follows. Take a litre flask filled with coloured water. Fit to it a cork through which passes a glass tube with a narrow bore (Fig. 99). By pushing in the tube or pulling it out as necessary, the level of water in the tube may be made to stand at any desired height (say at *A*). Put a mark there. If now the flask is suddenly immersed in hot water, the level of water in the tube sinks at first. The reason for it is that the flask has expanded due to heat, but no heat has yet entered the liquid. As the liquid heats up with time the water in the tube rises and moves past the former mark *A*.

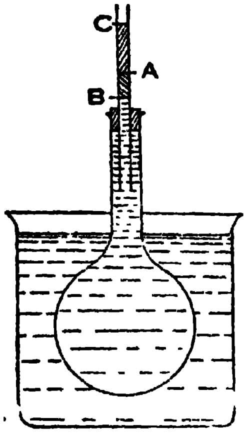


Fig. 99

Relation between real and apparent expansion. In Fig. 99 let *A* be the initial level of water in the tube at a temperature t_1, *B* the position which the level would occupy if the vessel alone expanded to temperature t_2 without any expansion of the liquid and *C* the final position of the level when the liquid is also at t_2.

The apparent expansion of the liquid is *AC* while its true expansion is *BC*; *AB* is the expansion in volume of the vessel, all for the same rise of temperature from t_1 to t_2.

From the figure $BC = AC + AB$

or *the true expansion* = *the apparent expansion* + *the expansion of the vessel*

Coefficient of real and apparent expansion. From the above it is clear that in the case of liquids we shall have two coefficients to consider, viz., (i) the *coefficient of real* (or *absolute*) *expansion*, which is defined as the true increase in volume per unit volume per unit rise of temperature, and (ii) *the coefficient of apparent expansion*, which is the observed increase in volume per unit volume per unit rise of temperature (measured in a vessel the expansion of which is disregarded).

If V_1= volume of a mass of liquid at t_1°C,

V_2= volume of the same mass of liquid at t_2°C,

then the mean coefficient γ between t_1° and t_2° is defined by the relation

$$\gamma = \frac{V_2 - V_1}{V_1 (t_2 - t_1)}$$

$$\text{or } V_2 = V_1 \{1 + \gamma (t_2 - t_1)\}.$$

Approximate relation between real and apparent coefficients of expansion

We know

real expansion = apparent expansion + expansion of the vessel.

If the original volume is V_1 and the rise in temperature t°C, then dividing the above relation throughout by $V_1 t$, we get *coefficient of real expansion* (γ_r)= *coefficient of apparent expansion* (γ_a)+*coefficient of volume expansion of the vessel* (γ_g)

$$\text{or} \quad \gamma_r = \gamma_a + \gamma_g$$

The following table provides the values of mean γ_r of some liquids over a range around 18°C unless otherwise specified. The values are per °C.

Substance	γ_r	Substance	γ_r
Water (5° to 10°C)	$5{\cdot}3\times10^{-5}$	Mercury	$18{\cdot}2\times10^{-5}$
,, (10° to 20°C)	$15{\cdot}0\times10^{-5}$	Glycerine	53×10^{-5}
,, (20° to 40°C)	$30{\cdot}2\times10^{-5}$	Benzene	124×10^{-5}
,, (40° to 60°C)	$45{\cdot}8\times10^{-5}$	Sulphuric acid (pure)	57×10^{-5}
,, (60° to 80°C)	$58{\cdot}7\times10^{-5}$	Olive oil	70×10^{-5}
Chloroform	126×10^{-5}	Alcohol (Ethyl)	110×10^{-5}
Pentane	159×10^{-5}	Turpentine	94×10^{-5}
Ether	163×10^{-5}		

99. Change of Density of a Liquid on Heating

Though expansion of the vessel masks expansion of the liquid, it cannot affect the change in density that a liquid undergoes with rise in temperature. The coefficient of real expansion bears a simple relation to the change in density and may be determined from observations on the latter.

Let V_0 = volume of a given mass m of a liquid at the lower temperature,
V = its *true* volume at a temperature $t°C$ *higher* than the colder temperature,
ρ_0 = density of the liquid at the lower temperature,
ρ = its density at the higher temperature,
and γ = the *mean coefficient of real expansion* between the two temperatures.

Then $\rho_0 = \frac{m}{V_0}$, $\rho = \frac{m}{V}$ and $V = V_0(1+\gamma t)$;

so $\rho = \frac{m}{V_0(1+\gamma t)} = \frac{m}{V_0} \cdot \frac{1}{1+\gamma t}$;

or $\rho = \frac{\rho_0}{1+\gamma t}$

When γt is small enough we may write

$$\rho=\rho_0(1+\gamma t)^{-1}=\rho_0(1-\gamma t)$$

100. Anomalous Expansion of Water

A liquid expands when heated; but water at $0°C$ is an exception. *When heated from $0°C$ water contracts as the temperature rises. This continues to about $4°C$*, above which point a rise in temperature causes expansion. Water, therefore, has maximum density at about $4°C$. An exact knowledge of the temperature is important, since the litre has been defined as the volume of one kilogram of water at its maximum density.

The anomalous expansion of water may be demonstrated by means of a simple experiment. Fig. 100 represents a glass vessel with a tube of narrow bore fitted to it. The tube carries a scale. The vessel contains mercury which occupies one-seventh part of the volume of the vessel. The rest is filled with water, which rises to a distance in the tube.

Since the cubical expansion of mercury is seven times that of glass, the inner volume of the vessel (not occupied by mercury) will not change when the vessel is heated. It is kept in melting ice for some time, when it acquires the temperature

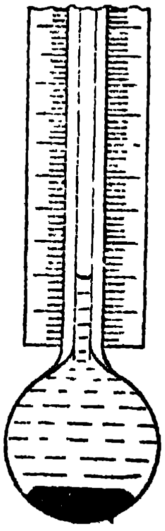


Fig. 100

0°C. The position of the water column in the tube is noted. If the vessel is now slowly heated it will be found that the water column descends at first, showing that water at 0°C contracts on heating. This goes on till 4°C is reached. Beyond that temperature the column ascends.

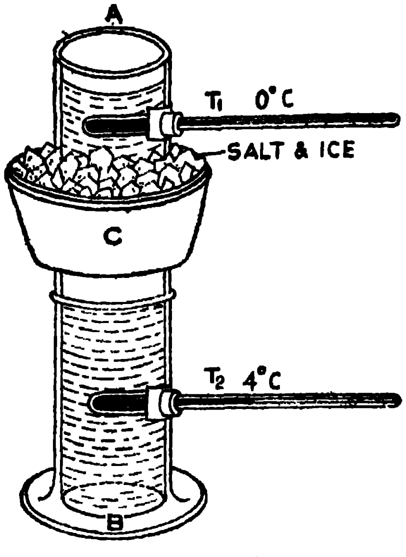


Fig. 101

Hope's experiment. The anomalous expansion of water may be demonstrated and the temperature at which the density of water is maximum may be determined by a simple experiment due to Hope. A tall glass vessel AB (Fig. 101) full of water has a freezing bath C of ice and salt round about its middle. Two thermometers T_1 and T_2 are fitted into AB near about the top and the bottom. Their readings are noted before the freezing mixture is applied. As the freezing mixture cools the water in its neighbourhood, the water becomes cooler and denser and sinks to the bottom. The temperature recorded by the lower thermometer T_2 gradually falls until it reaches about 4°C. By this time the upper thermometer T_1 has shown practically no change. While the reading of T_2 remains steady at 4°C, that of T_1 begins to fall slowly at first, and then more rapidly until 0°C is reached. This shows that water is densest at about 4°C. Careful measurements show that the temperature at which the density of water is maximum is 3·98°C. The manner in which the two thermometers change with time after the freezing mixture has been applied, is represented in Fig. 102.

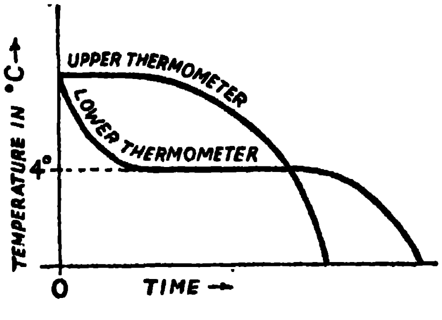


Fig. 102

Some consequences of the anomalous expansion of water. As water cools below 4°C it expands and becomes lighter. When ponds, lakes, rivers and seas in very cold countries freeze in winter, marine life is saved from extinction by the anomalous expansion of water. Figures 103 and 104 show how it is possible for fish and other aquatic animals to live in a lake when the temperature outside is 0°C or less. As atmospheric temperature comes down the upper layers of the lake cool, contract and sink to the bottom (Fig. 103). This goes on until the whole of the lake

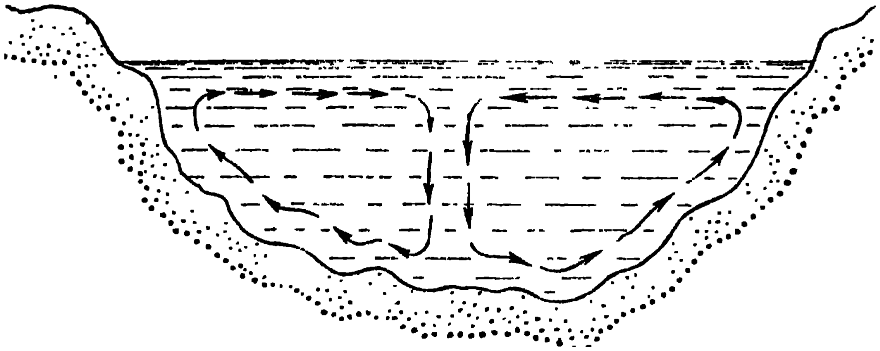


Fig. 103

acquires the temperature of 4°*C*. When the top layer's cool further, the cooler water does not sink as it is lighter than the water below. With further cooling the top layers gradually freeze (Fig. 104). Both ice and water are bad conductors of heat. So

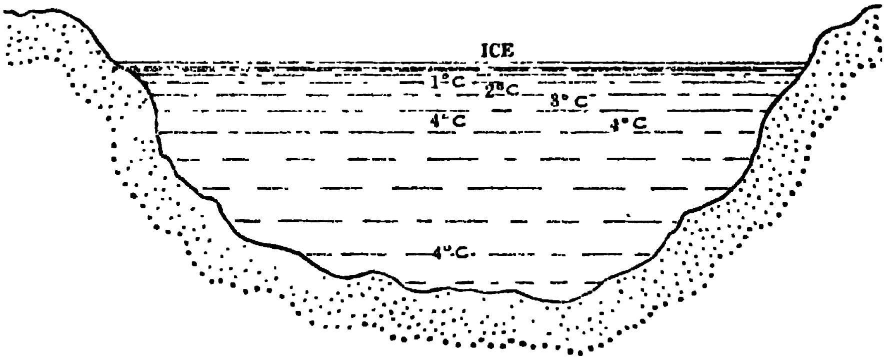


Fig. 104

the lower layers are protected to a great extent against freezing by the upper layers of ice and cold water. The result is that water continues to exist at the bottom though a thickness of ice may have formed at the top. This enables marine life to continue.

Exercises

1. Distinguish between absolute and apparent expansions of a liquid, and find the relation between the corresponding coefficients.

2. How does the density of a liquid change with temperature ? What is the peculiarity in the behaviour of water in this respect ?

3. How would you show that water at 0°*C* expands on heating and that water has its maximum density at 4°*C* ?

What is the consequence of this anomalous behaviour of water on marine life ?

4. A mercury in glass thermometer has a bulb 1 c.c. in capacity. It is desired to make each centigrade degree on the scale 5 mm long. Calculate the cross sectional area of the capillary tube. The coefficient of apparent expansion of mercury in glass is ·00016 per °*C* (*Ans* : ·0032 sq. cm)

5. What should be the volume of mercury to be put in a glass flask of capacity 720 c.c. so that the volume above the mercury remains the same at all temperatures. The coefficient of real expansion of mercury is ·00018 per °*C* and the coefficient of cubical expansion of glass is ·000025 per °*C* (*Ans* : 100 c.c.).

CHAPTER 13

EXPANSION OF GASES

101. Gases expand much more than solids or liquids for a given rise of temperature. Moreover a gas expands when its pressure is reduced. In determining the effect of temperature on the volume of a given mass of gas, it is therefore necessary to keep the pressure of the gas constant. Such a restriction is not necessary for solids or liquids, as a change of pressure causes but an inappreciable change of volume in them.

In discussing the expansion of gases it should be remembered that the **volume V of a given mass of a gas depends both on its temperature t and pressure P.** When the temperature is constant the relation between V and P is given by Boyle's law (§50).

To show that a gas expands when heated. Fit up a round bottomed flask with a rubber cork through which one end of a narrow glass tube projects a little into the flask. Invert the flask and introduce the other end into a beaker of water as in Fig. 105. On gently heating the flask with a flame, air bubbles will be seen to come out. The flame heats up the air inside the flask causing expansion of air within the flask, which escapes in the form of bubbles. On allowing the flask to cool the air contracts and some water is sucked up into the tube.

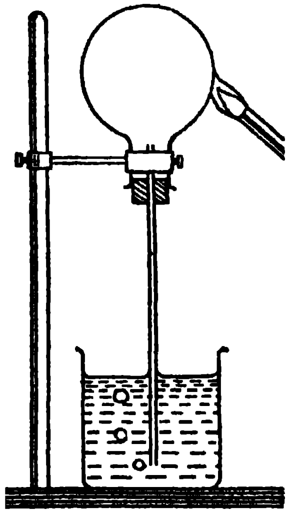


Fig. 105

102. Charles' Law

The relation between volume and temperature when the pressure is kept constant is known as Charles' Law. The coefficient (α_p) of increase in volume at constant pressure, is defined as the fraction of the volume at $0°C$ by which the volume of a given mass of gas increases for a rise in temperature of $1°C$, the pressure remaining constant. If V_0=volume of a fixed mass of gas at $0°C$, and V = its volume at $t°C$,

then
$$\alpha_p = \frac{V - V_0}{V_0 t},$$
which gives
$$V = V_0(1 + \alpha_p t).$$

Charles showed that *the coefficient of increase in volume of all* *gases at constant pressure is the same.* This statement is referred to as Charles' or Gay-Lussac's law.

According to Regnault, who measured the coefficient accurately,

$$\alpha_p = \frac{1}{273} \text{ per } {}^\circ C \text{ or } \cdot 00366 \text{ per } {}^\circ C.$$

In the light of Regnault's value for α_p **Charles' law** may be restated as follows :

For a fixed mass of any gas heated at constant pressure the volume increases by $\frac{1}{273}$ of the volume at 0°C for each centigrade degree rise in temperature.

103. The Absolute Scale of Temperature

Since for a gas $\alpha_p = \frac{1}{273}$ per °C, its volume V_t at any temperature t°C is given by

$$V_t = V_0 \left(1 + \frac{t}{273}\right)$$

Thus the volume at 1°C is $V_1 = V_0 \left(1 + \frac{1}{273}\right)$,

the volume at −30°C is $V_{-30} = V_0\left(1 - \frac{30}{273}\right)$ etc.

If α_p remains the same whatever the temperature, then at a temperature of −273°C,

$$V = V_0\left(1 - \frac{273}{273}\right) = 0$$

At a temperature below −273°C the volume would be *negative.* Since we cannot conceive of a negative volume, temperatures lower than −273°C have no sense to us. We therefore take the temperature at which the *calculated* volume of a gas becomes zero, to be the lowest temperature possible and call this temperature the **absolute zero**. Any scale of temperature measured from the absolute zero is called an **absolute scale of temperature.** On the centigrade scale, the absolute zero will be taken as −273°C.

The absolute scale of temperature in which there are 100 degrees between the ice point (0°C) and the steam point (100°C), i.e., the absolute scale in which each degree *interval* is equal to a degree interval on the centigrade scale, is called the **absolute centigrade scale**. The readings on this scale are indicated as *degrees Kelvin*, or in symbols, °K. It should be noted that

a difference of temperature of 1°K

=a difference of temperature of 1°C

Since absolute zero on the centigrade scale is $-273°C$, the ice point on the absolute centigrade scale is $273°K$ and any temperature $t°C$ is $(273+t)°K$, i.e.

$$t°C=(273+t)°K$$

104. Another form of Charles' Law

From Charles' law, we have for a given mass of gas at constant pressure,

$$V=V_0\left(1+\frac{t}{273}\right)=V_0\frac{273+t}{273} \qquad (104\cdot1)$$

where V is the volume of the given mass at t°C and V_0, that at 0°C.

Now $t°C=(t+273)°K=T°K$ (say),

while $0°C=273°K=T_0°K$ (say).

$$\therefore \text{ From Eq. } 104\cdot1, \quad V=\frac{V_0}{T_0}T$$

$$\text{or} \quad \frac{V}{T}=\frac{V_0}{T_0} \qquad (104\cdot2)$$

Since V_0 has a fixed value for the given mass, and T_0 is a constant, $\frac{V_0}{T_0}$ is also constant. We can therefore say that for a given mass of gas V/T is a constant, i.e., *at constant pressure the volume of a given mass of gas is directly proportional to its absolute temperature.* This assertion may also be taken as the statement of Charles' Law. In symbols, $V \propto T$ when P is constant.

105. The Ideal Gas Equation

A gas which obeys Boyle's Law and Charles' law strictly is called a *perfect or ideal gas.* Permanent gases (such as hydrogen, helium, oxygen and nitrogen) obey the laws with sufficient accuracy to be treated as perfect gases.

Let the pressure, volume and absolute temperature of a given mass of gas be P_1, V_1 and T_1.

(i) Keeping the pressure P_1 constant, let the absolute temperature be changed to T_2, and the new volume to V'. According to Charles' law $V'=V_1T_2/T_1$

(ii) Keeping the temperature T_2 constant, let the pressure be changed to P_2. Then, according to Boyle's law, the new volume V_2 is such that

$$P_2V_2=P_1V'=\frac{P_1V_1}{T_1}T_2$$

$$\text{or} \quad \frac{P_1V_1}{T_1}=\frac{P_2V_2}{T_2}=\text{a constant} \qquad (105\cdot1)$$

This is the ***ideal gas equation.***

If the volume remains constant, it follows that

$$\frac{P_1}{T_1}=\frac{P_2}{T_2}=\text{a constant,} \qquad (105{\cdot}2)$$

i.e., *the pressure of a gas at constant volume is directly proportional to its absolute temperature.*

106. Increase of Pressure with Rise of Temperature at Constant Volume

Equation 105·2 shows that when a given mass of gas at constant volume is heated, its pressure increases. Experiment shows that *at constant volume the pressure of a given mass of gas increases by a constant fraction of its pressure at 0°C for each degree centigrade rise of temperature.* This fraction is called the **pressure coefficient** (α_v) of the gas *at constant volume.*

If P_0 = pressure of a given mass of gas at 0°C,
P = pressure at temperature t°C,

then the fractional increase of pressure $=\frac{P-P_0}{P_0}$, and the fractional increase of pressure per °C rise of temperature (*i.e.*, α_v)

$$=\frac{P-P_0}{P_0 t}$$

or

$$P=P_0(1+\alpha_v t)$$

It is further found from experiment that α_v *is approximately equal to* $\frac{1}{273}$ *per °C for practically all gases.*

Exercises

1. State Charles' law and show how it gives rise to the idea of an absolute zero of temperature. What is an absolute scale of temperature ?

2.(a) A certain mass of gas is at 40°C. Calculate the temperature at which the volume will be doubled, pressure remaining the same.

(b) In the above example, if the volume is kept constant, what will be the temperature if the pressure increases to three times its initial value ? (*Ans* : 353°C ; 666°C)

3. Write down the ideal gas equation and clearly state the meaning of each symbol.

4.(a) A gas occupies 50 c.c. at 67°C and pressure 70 cm. of mercury. If the volume becomes 40 c.c. at 17°C, what is the new pressure ?

(b) A gas occupies 125 c.c. at 60°*C* and 75 cm of mercury. If the pressure be increased to 80 cm at temperature 30°*C*, what is the new volume ? (*Ans* : 74·6 cm of mercury ; 106·5 c.c.)

5. A thin glass bulb is sealed at 27°*C*, the inside pressure being 1 atmosphere. The maximum internal pressure the bulb can withstand is 95 cm of mercury. At what temperature will the bulb burst ? (*Ans* : 102°*C*.)

6. A balloon is filled with 610 cft of air and closed at 32°*C*. During the night the temperature falls to 25°*C*. Calculate the extra volume of air the balloon can now accommodate. (*Ans* : 14 cft)

7. A bicycle tyre contains air at a pressure of 2 atmospheres at 30°*C*. If the temperature rises to 40°*C*, what should be the pressure of air in the tyre, the volume remaining constant. (*Ans* : 2·07 atmos.)

8. At constant pressure 1 litre of a certain gas expands by 128 c.c. when heated from 0°*C* to 35°*C*. Calculate from these data the value of the absolute zero. (*Ans* : –273°*C*.)

CHAPTER 14

CALORIMETRY

107. Heat is a Measurable Quantity

In this chapter we shall consider how to measure heat and a few other thermal quantities. Such measurements form the subject matter of *calorimetry*.

Units of heat. Before we can measure a quantity it is necessary to fix upon a unit in terms of which it can be measured. The unit of heat is defined as the quantity of heat required to raise the temperature of unit mass of water by unity. This has given rise to three different units, viz.,

(i) the **calorie**, which is the unit of heat in the c.g.s. system. *It is the quantity of heat required to raise the temperature of one gram of water one degree centigrade.*

(ii) the **British Thermal unit** (abbreviated B.Th. U.), which is the unit of heat in the f.p.s. system. It is defined as the *heat required to raise the temperature of one pound of water one degree Fahrenheit.*

A commercial unit of heat is the *therm.*

1 Therm = 100,000 B.Th.U.

(iii) the **pound-degree-centigrade unit** (abbreviated lb. $°C$), which *raises the temperature of one pound of water* $1°C$.

It is clear from the definitions that the heat required to raise m grams of water through $t°C$ is $m \times t$ calories, while H calories will raise the temperature of m grams of water by H/m $°C$. We are thus led to the following relations between the units :

$$1 \text{ B.Th.U} = 1 \text{ lb} \times 1°F = 453{\cdot}6 \text{ gm} \times \tfrac{5}{9}°C = 252 \text{ calories},$$

$$1 \text{ lb } °C = 1 \text{ lb} \times 1°C = 453{\cdot}6 \text{ gm} \times 1°C = 453{\cdot}6 \text{ calories}.$$

108. Specific Heat

There are two definitions of specific heat in general use.

(i) The specific heat is the quantity of heat required to raise the temperature of unit mass of a substance by one degree. In the c.g.s. system, *the specific heat of a substance is the heat in calories required to raise the temperature of* 1 *gm of the substance through* $1°C$. According to this definition the unit is *calories per gm per* $°C$. Similarly, in the *fps* system the unit is *B.Th.U. per lb per* $°F$.

(ii) Specific heat is alternatively defined as the ratio of the heat required to raise the temperature of a given mass of the substance through any range to the heat required to raise the temperature of an equal mass of water through the same range. Or, to be more precise,

$$\text{Specific heat} = \frac{\text{heat to raise the temp. of unit mass of substance by unity}}{\text{heat to raise the temp. of unit mass of water by unity}}$$

As the ratio of two similar quantities is a pure number, specific heat, according to this definition, is a pure number. In the c.g.s. system the denominator is evidently 1 calorie. Thus if s be the specific heat of a substance,

the heat required to raise 1 gm. of the substance by $1\,°C$
$= s \times 1$ calorie $= s$ calories.

Similarly in the f.p.s. system, it is s B.Th.U. of heat.

We shall generally use c.g.s. units. From both definitions of specific heat it follows that if s be the specific heat of a substance,

heat required to change the temp. of 1 gm of a substance by $1\,°C = s$ calories;

heat required to change the temp. of m gm of a substance by $1\,°C = ms$ calories;

heat (H) required to change the temp. of m gm of a substance by $t\,°C = mst$ calories.

In symbols, $\quad$ **H = mst** $\quad$ (108·1)

Heat gained = mass × specific heat × temperature rise

Heat lost = mass × specific heat × temperature fall.

Table. Specific heats.
(in calories per gram per °C)

Substance	Sp. heat	Substance	Sp. heat	Substance	Sp. heat
Aluminium	·210	Nickel	·109	Glass	·12—·19
Copper	·091	Platinum	·032	Ice	·502
Gold	·030	Silver	·056	Marble	·22
Iron	·105	Tin	·054	Castor oil	·508
Lead	·030	Zinc	·092	Olive oil	·47
Mercury	·033	Brass	·088	Turpentine	·42

109. Thermal Capacity and Water-Equivalent

The **thermal capacity** *of a body is defined as the heat required to raise the temperature of the body by 1 °C.* If m be the mass of the body in grams and s the specific heat of its material (in cal/gm°C) it requires $H = mst$ calories for a rise of $t\,°C$. Then its thermal capacity C = heat required for $1\,°C$ rise of temperature $= H/t = ms$. In symbols,

$$C = ms \text{ cal}/°C \qquad (109{\cdot}1)$$

It is clear that the specific heat of a substance is its *thermal capacity per unit mass.*

The **water equivalent** *of a body is defined as the mass of water in grams which will be heated through* $1\,^\circ C$ *by the heat that raises the temperature of the body itself by* $1\,^\circ C$. If m is the mass of the body in grams and s its specific heat, the body requires ms calories in order that it may be heated by $1^\circ C$. Since 1 calorie heats 1 gm of water by $1^\circ C$, ms calories will heat ms grams of water by $1^\circ C$. Therefore, the water equivalent W of the body is ms grams, or in symbols,

$$W = ms \text{ gm} \qquad (109{\cdot}2)$$

From equations 109·1 and 109·2 it is clear that the numerical values of the thermal capacity and the water equivalent of a body are the same, but they are expressed in *different units*.

When the water equivalent W of a body is known, it follows from the definition of this quantity that the heat H which the body requires for a rise of temperature of $t^\circ C$ is

$$H = Wt$$

110. The Fundamental Principle of Calorimetry

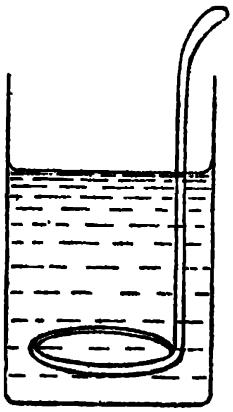


Fig. 106

In most experiments on calorimetry bodies at different temperatures are brought into close contact inside a calorimeter. It is a cylindrical copper vessel (Fig. 106) containing a liquid and a stirrer for stirring the liquid. Heat flows from the hotter to the colder bodies until all of them acquire a common temperature. The method is known as the method of mixtures.

Let us assume that no heat enters or leaves the calorimeter after the bodies have been brought into contact and that no chemical action takes place between the bodies. Then, from the principle of conservation of energy, we may say that,

Heat lost by the warmer bodies = heat gained by the colder bodies.

This is the fundamental principle of calorimetry.

The assumption that no heat enters or leaves the calorimeter must be fulfilled in calorimetric experiments. The calorimeter must be designed in such a way as to eliminate loss or gain of heat by conduction, convection and radiation. Precautions such as prolonged heating of the solid so that it may attain a steady temperature, quick transference of the hot body to the calorimeter, continuous stirring of the liquid and screening of the calorimeter from other sources of heat, must be taken.

Water is not suitable as a calorimetric liquid. Its specific heat is much higher than that of any other liquid. So the rise in temperature of water for a given supply of heat will be much less than that of other liquids. The accuracy with which a temperature difference can be measured diminishes as this difference itself becomes smaller. Therefore, with water as the calorimetric liquid the result is likely to be less accurate than if some other liquid, such as an oil, were used.

111. Some Calorimetric Measurements

Experiment 29. Determination of the specific heat of a solid. The body whose specific heat is to be determined is first of all weighed in air. It is then hung inside a steam chamber or heater as in fig. 107. The heater is a double-walled cylindrical chamber in which steam is admitted into the space between the walls. A thermometer records the temperature attained by the body. The bulb and the body must be close together and at the centre of the chamber. The lower end of the inner cylinder may be closed by a lid. The outer cylinder is covered with felt to prevent loss of heat.

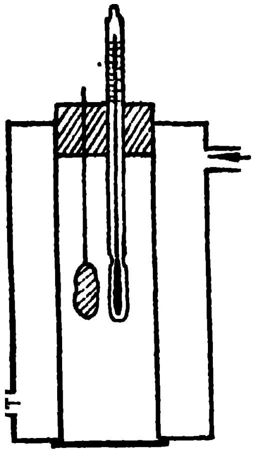


Fig. 107

Steam from a boiler enters the heater and heats the inner chamber. The temperature of the body rises and finally becomes steady at a value close to that of steam. Sufficient time should be allowed to enable the body to reach the steady temperature, which is recorded by the thermometer.

A known quantity of water is taken in a calorimeter whose water equivalent is known. The temperature of the calorimeter and its contents is ascertained.

The body is then quickly transferred into the calorimeter and the water in it stirred continuously till the contents of the calorimeter reach a common temperature. This value, which is the maximum temperature recorded by the thermometer in contact with water, is noted. Let—

the mass of the solid $= m$ gm,
its specific heat $= s$,
mass of water in the calorimeter $= m'$ gm,
water-equivalent of the calorimeter $= W$ gm,
initial temperature of water $= t_1 °C$,
initial temperature of the hot body $= t_2 °C$,
final common temperature $= t °C$.

Here the body loses heat while the calorimeter and its contents gain heat.

Heat lost by the body $= ms\,(t_2-t)$;
Heat gained by water $= m'\,(t-t_1)$;
Heat gained by the calorimeter $= W\,(t-t_1)$;

$$\therefore\ ms\,(t_2-t)=(m'+W)\,(t-t_1) \qquad (111{\cdot}1)$$

From this equation s may be determined, since all other quantities are known.

Note that the water equivalent W of the calorimeter may be calculated from its mass m_1 and the specific heat s_1 of the material of which it is composed. This mass includes that of the stirrer.

$$W=m_1 s_1.$$

Experiment 30. Determination of a high temperature by a calorimeter. The experiment as described above serves to measure the temperature of a furnace. A suitable piece of metal is heated in the furnace and dropped into the calorimeter. If the specific heat of the metal is known t_2 is the only unknown quantity in equation 111·1.

Experiment 31. Determination of the specific heat of a liquid. A weighed quantity of the experimental liquid is taken in a calorimeter of known water equivalent. A solid of known specific heat is heated to a known temperature and dropped into the calorimeter. The final common temperature of the calorimeter and its contents is noted. The method is thus the same as for determining the specific heat of a solid.

Let the mass of the solid $= m$ gm,
its specific heat $= s$,
the mass of the liquid $= m'$ gm,
its specific heat $= s'$,
water equivalent of calorimeter $= W$ gm,
initial temperature of the liquid $= t_1{}^\circ C$,
final common temperature $= t^\circ C$,
initial temperature of the solid $= t_2{}^\circ C$.
Then, heat lost by the solid $= ms\,(t_2-t)$,
heat gained by the liquid $= m's'\,(t-t_1)$,
heat gained by the calorimeter $= W\,(t-t_1)$.

$$\therefore\ ms(t_2-t)=(m's'+W)\,(t-t_1).$$

All quantities in the equation except s' are known. So s' may be calculated.

Experiment 32. Determination of water-equivalent of a calorimeter. Weigh the clean, dry calorimeter. Pour some water at the room temperature into it and weigh again. Note the temperature of this water. Add to it a nearly equal quantity of warm water of known temperature. Stir well and note the final common temperature. Weigh again after the calorimeter has cooled to the room temperature.

Let mass of the cold water $= m_1$ gm,
mass of warm water added $= m_2$ gm,
initial temperature of calorimeter and its contents $= t_1°C$,
temperature of warm water $= t_2°C$,
final common temperature $= t°C$,
water equivalent of calorimeter $= W$gm.

Heat gained by calorimeter and cold water $= (m_1 + W)(t - t_1)$
Heat lost by warm water $= m_2(t_2 - t)$

$$\therefore \qquad (m_1 + W)(t - t_1) = m_2(t_2 - t) \qquad (111.2)$$

whence W may be determined.

Though the experiment is simple it is difficult to get an accurate result from it, m_1 and m_2 should be nearly equal masses, and just enough to fully cover the bulb of the thermomater. t_2 may be about 20°C above t_1. The thermometer should read at least one-fifth of a degree.

Even better results may be obtained by using an oil of known specific heat, S, in place of water. If m_1 and m_2 are the masses of the cold and warm oil respectively these quantities should be multiplied by S in Eq. 111·2.

112. Calorific Values of Fuels

The number of calories of heat given out when 1 gm of coal or any other combustible substance is burned is called its *calorific value*. The higher the calorific value of a particular kind of coal, the higher the price we ought to pay for it. As the heat produced varies considerably with the quality of the coal, samples are tested for their calorific values. A known mass is burned in a nickel crucible, the apparatus being so designed that the heat developed goes into the water contained in the surrounding vessel. From the rise of temperature of this water and its mass, the heat produced may be calculated. The calorific value of an average priced coal is about 8000 cal/gm. The price paid for coal gas is also based on its calorific value.

Calorific values of foodstuff are also determined in the same way. Foodstuffs are the fuel which the body burns to derive energy.

Exercises

1. Define the calorie and the British Thermal Unit and express the latter in terms of the former. What is a therm ?

2. Define specific heat and show that the heat lost or gained by a substance is equal to the product of its mass, specific heat and change in temperature.

3 (a) What is the fundamental principle of calorimetry ?

(b) Define thermal capacity and water equivalent. What are the units in which they are expressed ?

4. Fill up the gaps in the following table :

	Mass	*Specific heat*	*Change of temperature*	*Heat*
(a)	50 gm	·09	30°*C* to 90°*C*	...
(b)	10 lb	...	100°*F* to 212°*F*	1120 B.Th.U.
(c)	100 gm	·21	40°*C* to ...	1680 calories
(d)	...	·25	20°*F* to 200°*F*	225 B.Th.U.
(e)	20 kgm	·4	20°*C* to 30°*C*	...

Ans: (a) 270 calories, (b) 1, (c) 120°*C*, (d) 5 lb, (e) 80,000 cal.

5. An iron ball weighing 50 gms was placed in a furnace for sometime and then quickly transferred to a calorimeter containing 1 kgm of water at 35°*C*. It raised the temperature of the calorimeter and its contents to 40°*C*. The water equivalent of the calorimeter is 125 gm and the sp. heat of iron ·12. Calculate the temperature of the furnace. (*Ans* : 977·5 °*C*).

6. A metal ball at 130°*C* and weighing 20 gm is dropped into a calorimeter (water equivalent 10 gm) containing 50 gm of a liquid (specific heat ·5) at a temperature of 40°*C*. The final temperature is 50°*C*. Calculate the specific heat of the metal. (*Ans* : 0·22).

7. A calorimeter of water equivalent 25 gm contains 100 gm of an oil at 40°*C*. A solid of specific heat ·1, weighing 50 gm is heated to 120°*C* and quickly dropped into the calorimeter. The resulting temperature is 45°*C*. Calculate the specific heat of the oil. (*Ans* : 0·5).

8. A calorimeter contains 500 gm of water at 30°*C*. 200 gm of water at 90°*C* are poured into it. If the water equivalent of the calorimeter be 10 gm what is the resulting temperature ? (*Ans* : 46·9°*C*).

9. 200 gm of an oil of sp. heat ·42 are dropped at a temperature of 60°C into a calorimeter containing 36 gm of water at 23°C. The resulting common temperature is 30°C. What is the water equivalent of the calorimeter ? If the mass of the calorimeter be 100 gm what is the specific heat of the material ?

(*Ans* : 10 gm ; 0·1)

10. Which of the two following cases requires the greater quantity of heat ?

(i) 500 gm of water heated from 35°C to 95°C.

(ii) 4 lb of water heated from 100°F to 212°F.

(*Ans* : The second case).

11. Calculate the cost of heating 1 ton of water from 100°F to its boiling point with gas costing 0·75 nP. per therm. 1.Therm = 100,000 B.Th.U. (*Ans* : Rs. 2·03).

12. A room contains 60 litres of air (sp. heat ·24) at 30°C weighing 1·3 gm/litre. How much heat is required to raise the temperature of the air to 40°C ? (*Ans* : 187·2 calories).

13. When 1 lb of a certain fuel is burned, 10,000 B.Th.U. are evolved. What quantity of the fuel will be required to raise the temperature of 50 gallons of water (1 gallon weighs 10 lb) from 112°F to the boiling point ? (*Ans* : 5 lb).

14. On pouring 1 kgm of water at 90°C into a hot water bottle at 35°C, the temperature of the water falls to 85°C. Calculate the weight of the bottle. The specific heat of the material of the bottle is ·2. (*Ans* : 500 gm).

15. The calorific value of coal gas is 500 B.Th.U. per cft. What volume of the gas would be required to bring 10 lb of water at 92°F to the boiling point, if the efficiency of the heater is 40%. (*Ans* : 6 cft.).

16. A can placed on a flame absorbs heat at the rate of 200 calories per sec. How long will it take to raise the temperature of 3 kgm of water from 30°C to the boiling point assuming that the can alone absorbs 3000 calories of heat.

(*Ans* : 17 min. 45 secs.)

CHAPTER 15

CHANGE OF STATE: SOLID TO LIQUID

113. Melting and Freezing

When a substance changes from the solid to the liquid state the process is called *melting*. The reverse process, i.e., a transformation from the liquid to the solid state is called *solidification*, or *freezing*.

The temperature at which melting takes place under a pressure of one standard atmosphere is called the **normal melting point.** Similarly, the temperature at which solidification takes place under a pressure of one standard atmosphere is called the *temperature of solidification*, or **freezing point.**

For a pure crystalline substance these two temperatures are the same and a sharply defined one. Each substance has its own particular temperature at which it melts. The change from solid to liquid begins only when this temperature is reached. *The substance remains at this temperature until the change is complete.*

Non-crystalline substances such as fat, wax, glass etc., pass through an intermediate viscous state before melting. They do not possess a fixed or sharply defined melting point. In some of these substances melting and freezing do not take place at the same temperature. Thus butter melts between 28°*C* and 33°*C* but solidifies between 23°*C* and 20°*C*. Impure substances and mixtures do not as a rule have a sharp melting point. Melting point, therefore, is a test of purity.

The following table gives the melting points of a few substances :

Substance	Melting point	Substance	Melting point
Brass	800°*C* to 1000°*C*	Paraffin—Hard	52°*C* to 58°*C*
Carbon	about 3500°*C*	Soft	38°*C* to 52°*C*
Copper	1083°*C*	Platinum	1755°*C*
Gold	1063°*C*	Silver	960°*C*
Iron (cast)	1100°*C* to 1300°*C*	Sulphur	115°*C*
Lead	327°*C*	Tin	232°*C*
Mercury	—39°*C*	Tungsten	3000°*C* (approx.)
Naphthalene	80°*C*	Zinc	418°*C*

Melting point is generally lowered by the presence of other substances. Plumber's solder, which is an alloy of two parts by weight of lead and one part of tin, melts at 180°*C*, though lead melts at 327°*C* and tin at 232°*C*.

A fire-fighting and alarm device used in factories makes use of such low melting alloys. Water pipes are fixed to the ceiling, and at intervals nozzles filled with the alloy are fitted. If fire breaks out the alloy melts and water is sprayed into the room. Another part of the device rings a warning bell at the same time.

114. Determination of Melting Point

(*a*) **Experiment 33. By the capillary tube method.** To determine the melting point of a substance like paraffin, melt some paraffin in a vessel and dip one end of a capillary tube in it. Some liquid will be drawn in. When the liquid solidifies, seal one end of the tube carefully. Fasten a thermometer to the capillary tube so that the bulb and the paraffin in the tube remain side by side. Place the combination vertically in a beaker full of water (Fig. 108). Slowly heat the water and stir it. Record the temperature when the paraffin melts. Then remove the flame, but continue stirring. Note the temperature when the paraffin solidifies. The mean of the two temperatures gives the melting point.

If the melting point of the solid be higher than 100°*C* the liquid in the beaker should be such that its boiling point is higher than the melting point to be determined. Cotton-seed oil may be used upto a temperature of about 250°*C*.

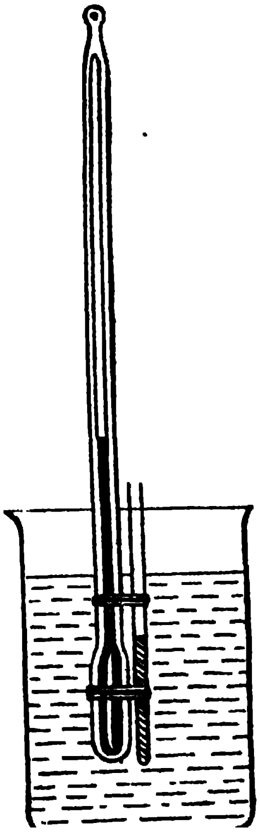


Fig. 108

(*b*) **Experiment 34. Graphical method.** Melt a quantity of the substance in a wide test tube and insert the bulb of a thermometer in the liquid. Now allow the liquid to cool *slowly* and record the temperature at intervals of half a minute, till the whole mass appears to have solidified. It will be seen that *the temperature remains constant during solidification.* If the time and temperature be now plotted on a graph paper, the general nature will be as shown in fig. 109. The portion of the curve parallel to the time axis and between the liquid and the solid states,

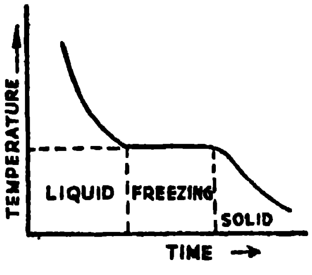


Fig. 109

indicates freezing and the corresponding temperature is the freezing point (or melting point, if the substance is crystalline).

115. Latent Heat and its Action

When heat is supplied to a substance its temperature rises. It is supposed that its molecules vibrate with greater and greater amplitude as the temperature rises. The energy required for the larger vibrations comes from the heat supplied. This heat is sometimes called *sensible heat.*

But the heat supplied to a solid during melting does not cause any rise in temperature. Where does then the energy go ? *The energy supplied goes to loosen the mutual attraction between the molecules, thus converting the solid into the liquid state.* This heat which has to be supplied to a solid at its melting point to bring about a change in state without change in temperature is called **latent heat.** The latent heat increases the potential energy of the liquid molecules, and is given out by the liquid when it freezes.

Quantitatively, **latent heat of fusion** (or **melting**) is the amount of heat required to convert one gram of solid to liquid without any change of temperature. The same amount of heat will be given out by 1 gram of the liquid when it solidifies at its freezing point without change in temperature.

Thus the statement that "Latent heat of ice is 80" means 80 calories of heat will be required to convert 1 gram of ice at 0°*C* to 1 gram of water at 0°*C* and, conversely, 80 calories of heat will be given out by 1 gram of water at 0°*C* to be converted into 1 gram of ice at 0°*C*.

Latent heat is expressed in heat units per unit of mass, i.e., in *calories per gram or B.Th.U per lb.*

The following table shows the melting points and latent heats of a few substances :—

Substance	Melting point (in °*C*)	Latent heat (in cal/gm)
Acetic acid	16·7	43
Ammonia	−75	108
Benzene	5·4	30
Glycerine	18	48
Ice	0	80
Lead	327	6
Mercury	−39	3
Sulphuric acid	10·3	24
Tin	232	14

The latent heat of water may be made to serve a useful purpose in preventing damage to meat, fruits and vegetables in storage. Extreme cold may freeze their sap or otherwise reduce their food value. To prevent such freezing a few buckets of water are kept in the store room. When the temperature of the room goes below 0°*C* the water in the buckets gradually freezes and gives out its latent heat. This raises the temperature of the room and it may not go down far enough to freeze the sap.

116. Determination of the Latent Heat of Ice

Experiment 35. Take a clean dry calorimeter with a wire gauze stirrer and weigh it empty. Fill it partly with water and weigh again. The difference gives the weight of water taken (m gm).

Put a thermometer in it and record the initial temperature. Let it be t_1°*C*.

Take one or two *small* pieces of ice and soak away all water from them by means of a piece of blotting paper. Holding the ice in the blotting paper, drop the former into the calorimeter. Stir the water keeping the ice below the stirrer. Note the lowest temperature reached. Let it be t_2°*C*.

Wait till the calorimeter and its contents are again at the room temperature. Weigh them again. The excess in weight gives the amount of ice added. Let the mass of ice added be m' gm.

Let W be the water equivalent of the calorimeter, and L, the latent heat of ice per gram.

Then, heat lost by the calorimeter $= W\,(t_1 - t)$;

heat lost by the water $= m\,(t_1 - t)$;

heat gained by ice on melting $= m'L$;

heat gained by water formed of ice $= m't$.

Since heat lost = heat gained

we have $(W+m)\,(t_1 - t) = m'L + m't$

$$\therefore\ L = \frac{(W+m)\,(t_1 - t)}{m'} - \;;$$

Note : To guard against such a fall in temperature of the calorimeter as will deposit water from the atmosphere on its sides, only one or two small pieces of ice should be added.

117. Change of Volume on Melting

The volume of most solids increases on melting ; but there are a few substances, viz., ice, cast iron and type metal, which show a decrease in volume on melting. This property of cast iron and type metal is taken advantage of in making castings of various designs from iron and in the manufacture of types for printing.

The fact that ice floats in water is of great consequence in keeping aquatic animals safe in the seas of the arctic regions or in waters which freeze in winter.

In expanding on solidification water exerts a considerable force, which way be exhibited as follows :

A flask made of iron is provided with a screw stopper. It is completely filled with water previously boiled to expel all dissolved air. The stopper is screwed down and the flask put in a freezing mixture for sometime. It will be found that the iron flask bursts as a result of the pressure exerted on it by the water during expansion on freezing.

If however the air in the flask or the dissolved air in water is not fully expelled, the vessel may not burst. The increase in volume may in that case be accommodated by the diminution of the volume of air under increased pressure.

In cold countries water pipes are sometimes found to burst in winter, hot water pipes bursting more often than cold water pipes. Hot water contains much less dissolved air than cold water. In cold water pipes the dissolved air is liberated on cooling and makes room for the increase of volume of water on solidification.

Boulders which have retained some water in their cracks may fall into pieces when, in cold winter nights, the water in the cracks freezes and expands.

118. Effect of Pressure on Melting Point

Melting points of substances are slightly affected by pressure. Most liquids contract on cooling and this continues when they turn solid. In these cases, increase of pressure assists the change from liquid to solid. Such a liquid will therefore, freeze more easily, that is, at a higher temperature when under pressure. Liquids like water which expand on freezing, will freeze at a lower temperature when subjected to pressure. Here pressure hinders the change ; hence it becomes necessary to reach a lower temperature before freezing can take place.

Regelation. The familiar phenomenon of fusion of two pieces of ice to form a single piece by pressure is an example of lowering of melting point under pressure. When two pieces of ice are pressed together the points of contact between them are subjected to high pressure. Hence the melting point at these places is lowered. The actual temperature at these points, i.e., $0°C$, is higher than the melting point under the condition of high pressure. Hence ice melts at these points taking the latent heat from their immediate surroundings. When the pressure is released the water around the points of contact freezes as it is lower than 0°*C* due to the abstraction of latent heat from it during melting. The two pieces are thus cemented together. The melting of ice by pressure followed by solidification of the water so formed on removal of pressure is known as **regelation.**

Bottomley's experiment. A single turn of *bare, thin copper wire* is wound round a large block of ice supported on two wooden blocks as shown in fig. 110 and is loaded by a weight of several pounds. The copper wire will be seen to penetrate through the block of ice and pass out of it, while the block remains intact. Due to the load the thin wire exerts a rather high pressure on the ice just below it and lowers the melting point at that place. The ice at that place, therefore, melts taking the latent heat from the surroundings through the copper wire. The water so formed flows round the wire which sinks, and freezes again due to the removal of the pressure on it. The latent heat it gives out passes through the wire and supplies the latent heat of fusion for the ice below the wire. This process is repeated and each layer is cemented no sooner the wire cuts through it.

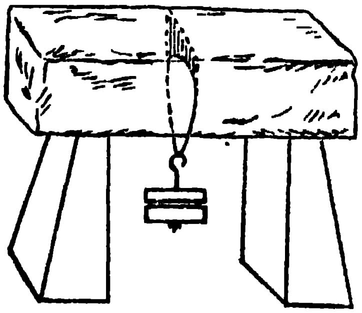


Fig. 110

For the success of the experiment it is necessary that heat be allowed to flow from the solidifying water to the ice to be melted below the wire. The wire must, therefore, be made of a material which conducts heat well. The experiment will not succeed if the wire is a bad conductor of heat.

For an increase in pressure of one atmosphere the melting point of ice diminishes by ·0075°*C*. The pressure should be about 1850 lb. wt. (i.e., 22 maunds) per square inch so as to make ice melt at $-1°C$.

119. Freezing Point of a Solution

The freezing point of a solution is always lower than that of the pure solvent. But when a solution begins to freeze the pure solvent separates out as crystals and the solution becomes more concentrated. This behaviour is often made use of in cold countries to obtain salt from sea-water. Sea-water begins to freeze at about $-2°C$. As water separates out by freezing, the solution gains in salt. It is finally evaporated by heat.

In cold weather water in the radiators of cars may freeze and burst the tubes. To prevent it glycol or glycerine is mixed with the water. The freezing point is lowered thereby.

Presence of dissolved substances in liquids always lowers the freezing point.

Exercises

1. What is latent heat? What does it do? What do you understand by the statement that latent heat of ice is 80?

Describe an experiment for measuring the latent heat of ice.

2. How is melting point affected by pressure? What is regelation? Explain how a loaded copper wire can pass through a block of ice. Will the experiment succeed if a silk thread replaced the copper wire? Explain your answer.

3. Describe an experiment to show that water exerts a considerable force while freezing. What effect has it on weathering of rocks?

4. 3 lb. of copper heated to 272°*F* are placed in a block of ice. How much of ice will melt? Sp. heat of copper is 0·1 and latent heat of fusion of ice 144 B.Th.U. per lb. (*Ans* : $\frac{1}{2}$ lb.)

5. A piece of ice at 0°*C* and weighing 15 gm is dropped into 85 gm of water at 20°*C* in a copper calorimeter. When all the ice melts the temperature falls to 10°*C*. If the latent heat of fusion of ice is 80 calories per gm, calculate the water equivalent of the calorimeter (*Ans* : 50 gms.)

6. 40 gm. of ice at −15°*C* (sp. heat 0·5) are mixed with 150 gm of water at 90°*C*. What is the resulting temperature?
(*Ans* : 45·25°*C*)

7. 250 gm of copper at 90°*C* are dropped into a copper calorimeter weighing 100 gm and containing 10 gm of ice and 25 gm of water at 0°*C*. The specific heat of copper is ·1 and latent heat of fusion of ice, 80 calories per gm. Find the resulting temperature. (*Ans* : 20·7°*C*.)

8. A 100 gm brass ball is cooled to the temperature of liquid air, i.e., −190°*C* and dropped into a calorimeter containing water at 0°C. Calculate the mass of ice formed (neglecting absorption of heat from the calorimeter). The specific heat of brass in this range is ·08 and latent heat of fusion of ice is 80 calories per gm.
(*Ans* : 19 gm.)

9. A quantity of ice at 0°*C* is added to 50 gm of water at 30°*C* in a calorimeter of water equivalent 10 gm. The final

common temperature is 10°*C*. Calculate the mass of ice added. Latent heat of fusion of ice is 80 calories per gram.

(*Ans* : $13\frac{1}{3}$ gm.)

10. A brass ball weighing 10 gm and heated to 250°*C* is put inside a cavity in a block of ice, of which 2·7 gm melt. If the latent heat of fusion of ice is 80 calories per gram, calculate the specific heat of brass. (*Ans* : ·09.)

11. A metal vessel containing 250 gm of water at 30°*C* is placed in a refrigerator which abstracts heat at the rate of 275 calories per minute. Calculate the time taken by the water to be converted into ice at 0°*C*. The latent heat of fusion of ice is 80 calories per gram. The weight of the vessel may be neglected.

(*Ans* : 1 hr. 40 min.)

12. 100 gm of ice at 0°*C* are added to 200 gm of water at 30°C. What will be the final temperature?

13. Suppose that the latent heat of ice were suddenly reduced to half its value. What may be the effect of such a change on the snows of the Himalayas? (Answer in ten lines.)

CHANGE OF STATE : LIQUID TO VAPOUR

120. Vaporization and Vapour Pressure

When a liquid is heated it changes into vapour. Such change is called *vaporization*. It can take place in two ways, viz., (i) by *evaporation*, and (ii) by *boiling*. The reverse process of change of state from vapour to liquid is called *condensation* or *liquefaction*. Before proceeding to consider the processes of evaporation and boiling in some detail it may be profitable for us to know that a vapour exerts pressure like a gas. This may be seen as follows.

Take a barometer tube filled with mercury and invert it over a mercury trough. Introduce a few drops of a liquid, such as water, alcohol or ether, into the tube with the help of a bent pipette (Fig. 111). As the first drops rise through the mercury column and reach the vacuous space above, they turn completely into vapour. The vapour so formed depresses the mercury column, showing that it exerts pressure like a gas.

As you introduce more and more of the liquid into the tube, a state will be reached when it no longer evaporates, but accumulates on the mercury surface. When this condition is reached, the pressure exerted by the vapour does not increase any more.

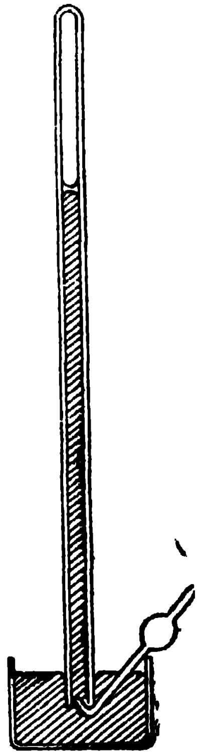


Fig. 111

These show that (i) a vapour exerts pressure, and (ii) there is a limit to the amount of vapour that a space (here the Torricellian vacuum) can hold. If we increase the temperature of the Torricellian space by surrounding it with a suitable bath, it will be found that the maximum vapour pressure exerted by the liquid increases with rise of temperature, and the space can hold more vapour at the higher temperature.

121. Evaporation

Evaporation is the slow process of conversion of a liquid into the vaporous state. It is easy to confirm the following facts about evaporation. Many of them are part of our daily experience.

(i) Evaporation takes place at all temperatures and under all pressures.

(ii) The rate of evaporation increases with rise of temperature.

(iii) The rate of evaporation also depends on the pressure. The lower the pressure on a liquid the faster will it evaporate. Evaporation is quickest in a vacuum.

(iv) The greater the area of the exposed surface the greater the evaporation.

(v) Evaporation is speeded up by a current of air over the liquid.

(vi) The rate of evaporation depends on the amount of the vapour of the liquid present over the liquid surface. The higher it is the slower is the evaporation. A moist cloth dries much slower on a wet day than on a dry day because there is more moisture in the air on the wet day.

(vii) Of two liquids the one which has a lower boiling point will evaporate more quickly. Ether evaporates much more quickly than alcohol, and alcohol quicker than water. The boiling points of ether, alcohol and water are 35°*C*, 78°*C* and 100°*C* respectively.

(viii) *Heat is required for evaporation. If it is not supplied the liquid cools itself or its surroundings to get the heat* (Also see § 127).

122. Boiling

A. Nature of boiling. Take some water in a flask fitted with a rubber stopper. Pass through it a thermometer, an open-tube manometer, and a delivery tube provided with an india rubber tubing and a pinch cock (Fig. 112).

Heat the flask slowly over a Bunsen flame. As the temperature gradually increases you will find more and more steam rising *from the surface.* (This is not boiling). Also, the dissolved air in the water forms into small bubbles, rises to the surface and escapes. At about 70°—80°*C*, small bubbles of water vapour will be seen to form at the bottom of the flask. They rise and collapse as they reach the upper layers of colder water. At this stage a simmering sound is heard. In the final stage the bubbles rise to

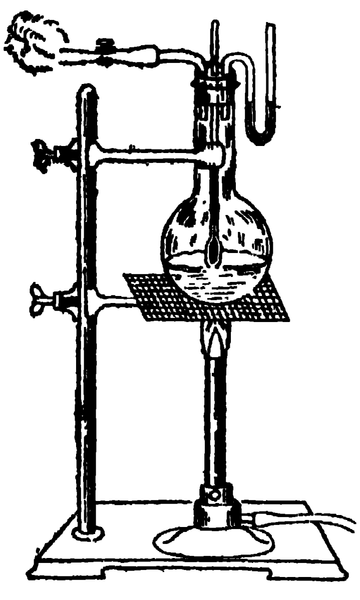


Fig. 112

the top and burst at the surface. When this happens we say that the liquid is boiling. *During boiling vapour rises to the surface from all points throughout the liquid. The temperature of the liquid remains constant so long as the boiling continues.*

The vapour within the flask is invisible like air. As it comes out of the delivery tube it condenses into small water globules which look like a white cloud. This cloud soon vanishes as the globules evaporate in the open air.

During boiling, the manometer registers a constant pressure. When the delivery tube is wide open, mercury in both limbs of the manometer stands at the same height, showing that the pressure inside the flask is equal to that of the atmosphere. If boiling has continued long enough we may safely conclude that air inside the flask has been completely replaced by water vapour. So the pressure inside the flask is that exerted by the water vapour only. This shows that **when there is a free communication with the atmosphere a liquid boils when its vapour pressure is equal to the atmospheric pressure.**

The constant temperature at which a liquid boils under normal atmospheric pressure is called its **normal boiling point**. The boiling point of water under a pressure of 76 cm of mercury is 100°*C*.

B. Effect of pressure on boiling point.

(1) **Increase of pressure on a liquid raises its boiling point.** If during brisk boiling we close the pinch cock partially, so that the flow of steam is appreciably impeded, the manometer registers a higher pressure in the flask. The thermometer also registers a higher temperature when the liquid boils under this condition of increased pressure on its surface.

We, therefore, conclude that an increase in pressure on the liquid surface results in an increase in the boiling point of a liquid.

(2) **Lowering of pressure on a liquid lowers its boiling point.**

(a) **Franklin's experiment.** Take a flask half full of water and boil it briskly. After boiling has proceeded for some time so that the air inside has been replaced by water vapour, remove the flask from the flame and quickly close it with a

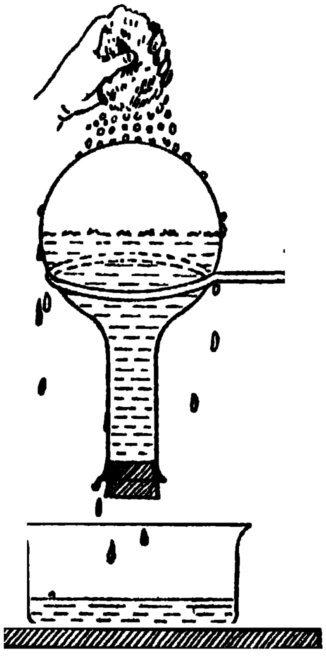


Fig. 113

rubber stopper. Note that boiling stops immediately. Now invert the flask on a retort holder and pour cold water on it (Fig. 113). The water within the flask will be seen to boil again. This boiling lasts for a short time. The experiment can be repeated several times after the first boiling.

The explanation is simple. Part of the water vapour above the surface of the liquid condenses into water due to the cooling of the surface of the flask. This results in a diminution of pressure on the liquid surface. If the pressure is lowered below the vapour pressure of the water in the flask, the liquid boils.

This shows that a liquid may be made to boil at a lower temperature than its normal boiling point. This happens when the pressure on it is reduced below the normal atmospheric pressure.

Boiling under reduced pressure has its application in industry. There are chemicals, such as H_2O_2, which decompose before reaching the normal boiling point. In manufacture, they are separated by distillation at a reduced pressure.

In the sugar industry, sugar is separated by crystallization from solution. The syrupy solution has a high boiling point. It is made to boil at a lower temperature by reducing the pressure on it. This economizes fuel consumption.

(b) **Boiling of water at room temperature.** Water may be made to boil at room temperature without heating. For this purpose a high speed exhaust pump and towers containing absorbents of water vapour, such as anhydrous $CaCl_2$, are necessary.

Take a rather small quantity of water in a flask and close it with a rubber stopper. Through the cork pass a thermometer and two glass tubes, one of which is connected to the pump and the other to a closed-tube manometer (Fig. 114). The pump is connected to the flask through two absorbing towers containing fused calcium chloride. See that the system is leak tight. Now work the pump. When the pressure falls considerably the water in the flask begins to boil at the room temperature. As the boiling is continued the thermometer may register a temperature a little lower than the room temperature.

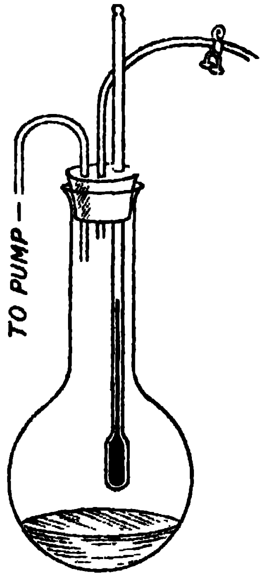


Fig. 114

C. Boiling point of a solution. Make an arrangement as in fig. 112. Add some salt to the water in the flask and put the thermometer bulb *within the liquid.* Note that boiling sets in at a higher temperature when the salt is added.

Slowly raise the thermometer above the liquid in the flask. Note that the temperature recorded gradually diminishes until it is the same as we would get if we had pure water boiling inside. This shows that the boiling point of a solution is higher than that of the pure solvent under the same pressure. But the temperature of the vapour is ultimately the same as for the pure solvent boiling under the same pressure.

D. Characteristics of boiling and factors affecting it.

Summarising, we may say as follows :

(i) Boiling takes place at a definite temperature under a definite pressure. The temperature at which a liquid boils under normal atmospheric pressure is called the **normal boiling point.**

(ii) For a pure liquid the temperature at which boiling takes place at a given pressure depends on the nature of the liquid.

(iii) *A liquid boils when its vapour pressure is equal to the pressure on its surface.* Hence *the boiling point of a liquid rises or falls with the pressure on it.*

(iv) At a given pressure, the boiling point of a solution is higher than that of the pure solvent. But the vapour formed has finally the same temperature in both cases.

(v) Boiling is accompanied by absorption of heat. The rate of boiling depends on the rate of heat supply.

Table. Normal boiling points of some substances in °*C*

Substance	Boiling point	Substance	Boiling point	Substance	Boiling point
Sulphur	444°	Water	100°	Ammonia	−34°
Mercury	357°	Alcohol	78°	Oxygen	−183°
Paraffin	350° – 530°	Chloroform	61°	Hydrogen	−253°
Glycerine	290°	Ether	35°	Helium	−269°

E. Difference between evaporation and boiling. From the characteristics of the two processes already discussed, we may say that the fundamental differences between them are the following :—

(i) Boiling takes place at a definite temperature under a definite pressure, but evaporation takes place at all temperatures under any pressure.

(ii) Boiling takes place throughout the bulk of the liquid while evaporation takes place only from the exposed surface.

123. Boiling at High Altitudes

Atmospheric pressure is greatest at the sea level and diminishes as we go up. Since the boiling point of a liquid diminishes with lowering of pressure, we should expect that in hill stations water will boil at a lower temperature than on the plains. This is actually so. On Mt. Blanc (15,782 ft.) water boils at 83°C ; on Mt. Everest (29,002 ft.) it would boil at 70°C. It has been estimated that at a height of 65,000 ft. water would boil at 37°C. At this height the water in our body, *if exposed*, would start boiling.

The following table gives the approximate temperatures at which water boils at the heights mentioned.

Height above sea-level (in ft.)	Boiling point (in °C)
2000	98
4000	96
7000	93
10,000	90
15,000	85
20,000	80
25,000	74

It would be seen that the boiling point of water diminishes roughly by 1°C per thousand feet (960 ft. to be more precise) of ascent above the sea-level upto about 20,000 ft. A knowledge of the boiling point of water can, therefore, give us an idea of the height of the place above sea-level.

124. Boiling under Increased Pressure

Cooking at temperatures less than 100°C is of great inconvenience. For cooking at high altitudes a special kind of cooker, called **pressure cooker** (the same thing as **Papin's digester**) is used. The principle of the cooker or Papin's digester may be explained with the aid of fig. 115. The vessel is made of thick sheet metal. The lid is screwed to it air-tight. A valve closes an opening in the lid. It is kept pressed in position by a suitable weight *W* acting through a lever. The weight and the position of *W* determine the pressure of the steam in the

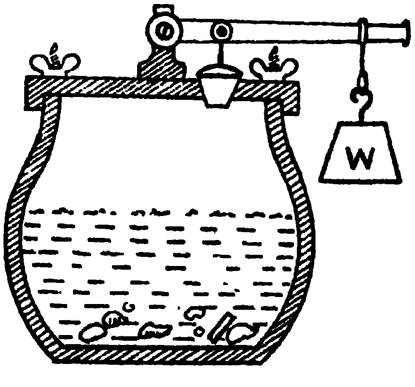


Fig. 115

cooker which will lift the valve. If the steam generated within the vessel attains a greater pressure, the valve will be forced open and the excess of pressure released. By suitably closing W the water inside may be made to boil at a desired temperature. If W exerts an extra pressure of about one atmosphere the boiling point of water on the plains will be about 120°C. Meat can be cooked in 20 minutes with such a cooker. The external force on the valve can also be applied by an adjustable spring.

Papin's digester is used in manufacturing paper pulp by boiling saw dust and caustic soda under pressure, in the manufacture of artificial silk and for extracting gelatine from bones. In boilers of steam engines steam is generated under high pressure by making water boil at a temperature higher than 100°C. Safety valves open when a pre-determined pressure is reached. Boilers of locomotive engines may have a steam pressure of about 250 lb. per sq. inch (i.e., over 15 atmospheres), the water boiling at about 200 °C.

In autoclaves and hospital sterilizers water may be made to boil at about 135°C corresponding to a pressure of 30 lb. per sq. inch above atmosphere pressure. At this temperature bacteria and bacterial spores are killed. Dressings, towels etc. are placed in the chambers of the sterilizer. High pressure steam circulates through an outer jacket. Canning food is cooking under pressure. Sealed cans are heated by high pressure steam to about 150°C to kill bacteria and also to cook. But it destroys the vitamins.

125. Latent Heat of Vaporization

We have seen that there is no change of temperature when a liquid boils under a constant pressure (§ 41 A). Nevertheless, heat must be supplied even at the boiling point in order that the liquid may boil. Heat is also necessary to evaporate a liquid. This heat, which does not raise the temperature of the liquid, but brings about a change of state from liquid to vapour, is called *latent heat of vaporization*.

There is some attraction between the molecules of a liquid, but in a vapour the mutual attraction between molecules is negligible. To convert a molecule of the liquid into a molecule of the vapour, it must be removed beyond the range of attraction of other liquid molecules. In other words, some energy is necessary to effect the change of state and this energy will not, as in the general case, go to raise the temperature of the liquid. Latent heat of vaporization supplies this energy. It may be looked upon as being stored up as potential energy of the vapour molecules, and is released as heat when the vapour condenses.

Quantitatively, **the latent heat of vaporiaztion is the amount of heat required to convert unit mass of a liquid into vapour at the same temperature.** It is measured in *heat units per unit mass*, generally in *calories per gram*. Latent heat is independent of the process by which the change of state is brought about. Whether the liquid changes into the vapour by evaporation or by boiling, it requires the same latent heat at a given temperature.

The statement that *the latent heat of steam is* 537 *calories per gram* means that when water at 100°*C* is converted into steam at 100°*C*, each gram will require 537 calories for the change of state only. When 1 *gm* of steam at 100°*C* condenses into water at 100°*C*, 537 calories of heat are released.

126. Determination of the Latent Heat of Steam

Experiment 36. Take a clean dry calorimeter and weigh it. Fill it more than half with water and weigh again. Note its initial temperature.

Boil water in a vessel *B* (Fig. 116) provided with a delivery tube leading to a steam trap *T*. From the trap only steam can enter into the calorimeter *C* through another tube. The condensed water will be arrested in the trap and can be drained off from time to time. A wooden screen separates the boiler *B* from the calorimeter *C*. Allow the steam to condense in the calorimeter for some time. Stir the water well and read off the final temperature after the supply of steam has been cut off.

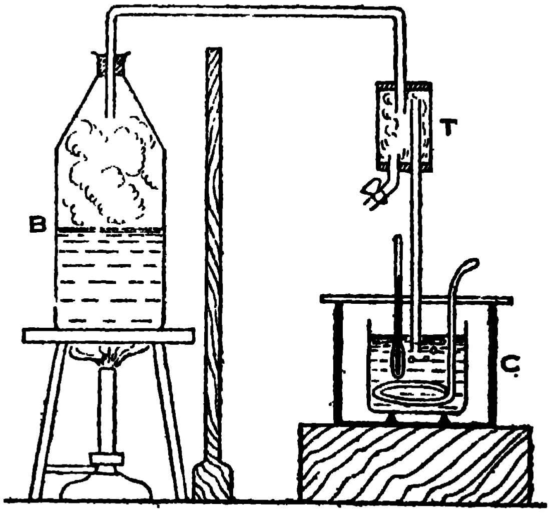


Fig. 116

Wait till the calorimeter and its contents reach the room temperature. Weigh them again. The difference from the second weight gives the mass of the steam condensed.

Let m gm = initial mass of water taken,
m' gm = mass of steam condensed,
W gm = water equivalent of the calorimeter,
t_1°C = initial temperature of water,
t°C = final temperature,
L = latent heat of steam in calories per gm,

Then, the heat lost by the steam at 100°C to condense into water at $100°C = m' \times L$ cal.

Heat lost by the water so formed to cool through $(100-t)°C$
$$= m' (100-t) \text{ cal.}$$

Heat gained by water and the calorimeter
$$= (m+W)(t-t_1) \text{ cal.}$$

$\therefore m'L + m'(100-t) = (m+W)(t-t_1)$, from which L can be calculated.

127. Cold caused by Evaporation

A liquid cannot change to vapour unless the requisite amount of latent heat is supplied. Hence when a liquid evaporates, the portion that evaporates draws its latent heat from its neighbourhood including the rest of the liquid if there is no external source of heat. As a result the liquid and its surroundings are cooled.

This explains a number of phenomena which we come across in every day life.

(i) When spirit is sprinkled on our skin, the spirit evaporates and takes the latent heat from the skin, which feels cold.

(ii) If after a bath we expose ourselves to a wind, or after perspiration sit below a fan the water on our body evaporates taking the latent heat from our body. We therefore feel cold.

In tropical countries an adult may perspire to the extent of a litre or more per day. A considerable part of it evaporates taking from the body the necessary latent heat, which is about 580,000 calories per litre. This is one of the ways our body gets rid of the superfluous heat.

(iii) Water in an earthenware pot is cooled in supplying the latent heat of evaporation to the water that comes out through the pores in the pot and evaporates.

(iv) The dentist sprays the gum with a liquid known as ethyl chloride before he makes an injection. The rapid evaporation of ethyl chloride cools and 'freezes' the gum so as to make it less sensitive.

Freezing by evaporation. (i) Pour a little water on a wooden block and place on it a thin tin can containing ether. Bubble air rapidly through the ether by means of a glass tube and hand bellows. The rapid evaporation of ether cools the water and causes it to freeze. The wooden block can then be lifted by the tin can.

(ii) Preparation of **dry ice,** which is solid carbon dioxide, is an example of freezing by evaporation. Liquid carbon dioxide contained in a cylinder (Fig. 117) issues in a jet through a valve. The rapid evaporation of the liquid cools it and a part freezes. This is collected in a muslin bag attached to the nozzle.

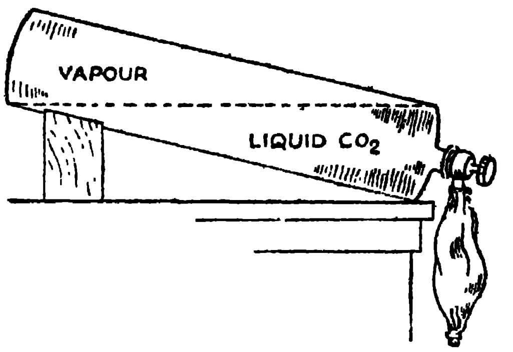


Fig. 117

Dry ice has a temperature of —78°C. It is used for cooling. The ice cream vendor uses dry ice to keep his goods from melting.

Refrigeration. Use of household refrigerators is on the increase. You often see them in a druggist store or a sweetmeat shop. They provide cold chambers, around 0°*C*, in which you can keep food, sera, vaccines etc. for a long time without deterioration. Large rooms can be kept cool in the same way as a household refrigerator and used for large scale storage of meat, fruit and vegetables for months. Ships which transport such items of food from one country to another are provided with cooling plants in their holds as in refrigerators.

Cooling in these devices is produced by the evaporation of a liquid under reduced pressure in a set of coils called the *evaporator coils*. The coolant may be ammonia, sulphur dioxide, ethyl chloride or freon (CF_2Cl_2).

Exercises

1. How would you show that a vapour exerts pressure and that the maximum pressure a vapour can exert depends on the temperature ?

2. State the principal facts about evaporation and boiling. What is the fundamental difference between them ?

3. How would you show that a liquid boils when its vapour pressure is equal to the pressure on it ?

4. What is the effect of change of pressure on boiling point ? What is meant by normal boiling point ?

Describe one experiment each to show the effect of (a) an increased pressure, (b) a reduced pressure on boiling point.

5. Mention some applications of boiling under (a) increased, (b) reduced pressure.

What broad inferences regarding height would you draw from the statement that water boils at 97°*C* at a certain hill station ? What would you expect the boiling point of water to be in a mine 2000 ft deep ?

6. What do you understand by the statement that 'the latent heat of steam is 540 calories per gram' ? Describe an experiment to measure this latent heat.

7. Give some examples of cold produced by evaporation. Mention some practical uses of it.

8. A can of water-equivalent 100 gm contains 500 gm of water at 40°*C*. Steam at 100°*C* is allowed to condense in the vessel. What quantity of steam is required to raise the water to its boiling point ? Latent heat of steam is 540 calories per gram. (*Ans* : $66\frac{2}{3}$ gm.)

9. What quantity of steam at 100°C should be condensed in a mixture of 300 gm of water and 30 gm of ice at 0°*C* so as to make the final temperature 30°*C* ? Latent heat of fusion of ice is 80 calories per gram and latent heat of steam, 540 calories per gram. (*Ans* : $11\frac{2}{3}$ gm.)

10. A calorimeter of which the heat capacity may be neglected, contains 1000 gm of water at 30°*C*. It is placed on an electric stove, when the temperature of water rises to 100°*C* in 10 minutes. Calculate the amount of heat absorbed by the water per minnte. How long will the water at 100°C take to boil away completely ? (*Ans* : 7000 calories/min ; $77\frac{1}{7}$ min)

11. By the evaporation of 25 gm of liquid ammonia 85 gm of ice at 0°*C* are formed from water at 20°*C*. If the latent heat of fusion of ice is 80 calories per gram, calculate the latent heat of vaporization of ammonia. (*Ans* : 340 calories per gram.)

CHAPTER 17

HYGROMETRY

128. Water Vapour in the Atmosphere

Evaporation is constantly taking place from the surface of water which covers more than two-thirds of the face of the earth. Many millions of tons of water are drawn up into the air every hour. Air must, therefore, always contain some water vapour. Though the quantity of water vapour is small compared with the other constituents of air, such natural phenomena as rain, dew, fog etc. are due to the condensation of moisture in air.

We feel comfortable when the moisture in air lies within certain limits. Good living conditions, therefore, require a controlled amount of moisture in the air. A similar control is necessary for carrying out some industrial processes connected with cotton and wool industries, artificial seasoning of timber, manufacture of artificial silk etc. Besides, forecasting of weather requires a knowledge of the amount of moisture in the air. It, therefore, becomes necessary for us to determine the moisture content in air at different places and times. *Hygrometry* is concerned with the measurement of the amount of moisture in the air.

129. Sautrated and Unsaturated Vapour

Before proceeding further it will be worth while to go back to § 120 where an experiment has been described to show that

(i) water vapour exerts pressure ;

(ii) there is a limit to the amount of water vapour that a space can hold ;

(iii) the pressure as well as the amount of vapour increases with rise of temperature.

When a space contains the maximum amount of water vapour it can hold at a given temperature, it is said to be *saturated with the vapour*. Otherwise, it is *unsaturated*. The terms saturated and unsaturated are also applied to the vapour itself. The maximum pressure which a vapour can exert at a given temperature is variously called **saturated vapour pressure** (abbreviated S.V.P. in this book), *maximum vapour pressure*, or simply, *vapour pressure* of the liquid at that temperature. It easily follows that at a given temperature, a saturated vapour exerts the maximum vapour pressure. If the pressure exerted by a vapour is less than the maximum it can exert at the temperature, the vapour is unsaturated.

130. Dew Point

Ordinarily air does not contain enough moisture to saturate it. If the air at any place is gradually cooled, it will tend towards saturation. At a certain temperature a given volume of air will be saturated by the moisture it contains. This temperature is called **dew point**. Further cooling causes dew to appear. When you add ice to water in a tumbler, dew forms on the outside of the tumbler due to the cooling of the surrounding air below the dew point.

The pressure due to the vapour actually present in the air is unaffected by this cooling. Since at the dew point the vapour is saturated, *the saturation vapour pressure corresponding to the dew point is also the actual pressure due to the vapour present.*

The S.V.P. of water vapour at different temperatures was determined by Regnault and compiled into a table known as Regnault's table. If the dew point at any place at a given time be determined, the actual pressure of vapour in the air could be found by reference to the table.

131. Humidity of Air

We often use the terms 'dry' or 'moist' with respect to air. The use of these terms is generally dictated by our *feeling* as to whether the air contains little or much moisture. In reality this feeling depends on two factors, viz., (*a*) the actual amount of moisture present in the air and (*b*) the moisture necessary to saturate the air under the existing conditions. Strictly speaking, our sensation of dryness or dampness of air depends on the ratio of (*a*) to (*b*). The humidity (or dampness) of air may be expressed in either of two ways, viz., (*i*) by the absolute humidity and (*ii*) by the relative humidity.

The **absolute humidity** is defined as the mass of water vapour present in unit volume of air. It is expressed in grams per cubic metre of air.

The **relative humidity** is the ratio of the actual mass of water vapour present in a certain volume of the air to the mass of water vapour required to saturate the same volume at the same temperature. It is usually expressed as a percentage. If 1 cu. metre of air at a particular temperature contains 18 gm of water vapour and if 30 gm of vapour are required to saturate it at the same temperature, the relative humidity is 60%.

Assuming that water vapour obeys Boyle's law up to the point of saturation, the density of the vapour, and hence the mass present in a given volume, is proportional to the pressure it exerts. This assumption is not quite correct, but the inaccuracy is not important for the purpose. This leads to another definition of relative humidity, viz.,

$$\text{Relative humidity} = \frac{\text{actual pressure of water vapour present in the air}}{\text{S.V.P. of water at the temperature of the air}}.$$

Since the actual pressure of the vapour present is equal to the saturation pressure at the dew point, we may also define relative humidity as

$$\textit{Relative humidity} = \frac{\textit{Saturation pressure of water at the dew point}}{\textit{Saturation pressure at the temperature of the air}}.$$

If the temperature of the air $= 35°C$, dew point $= 26°C$ S.V.P. at $35°C = 42{\cdot}02$ mm of mercury and S.V.P. at $26°C$ $= 25{\cdot}13$ mm of mercury,

$$\text{R.H.} = \frac{25{\cdot}13}{42{\cdot}02} = 0{\cdot}598 = 59{\cdot}8\%.$$

The rate of evaporation of water is determined by the relative humidity of air, and not by its absolute humidity. For comfort the relative humidity of the air in a room should be kept between 50 and 60.

132. Measurement of Humidity : Hygrometers

Hygrometers are instruments for measuring the humidity of air. There are different types, of which we shall describe two.

(i) Dew point hygrometer : Regnault's hygrometer.

The action of this class of hygrometers is based on the definition of relative humidity as the ratio of two pressures. They are used with Regnault's tables which give the S.V.P. of water at different temperatures. The hygrometer determines the dew point.

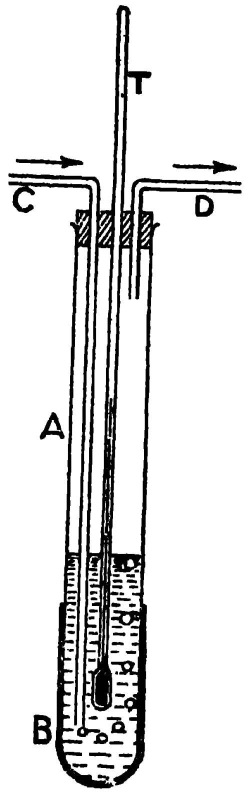


Fig. 118

A glass tube A (Fig. 118) closed at its upper end by a stopper carrying a thermometer T, is fitted at its lower end with a silver cap. Ether is poured into the cap and fills it completely. Air is sucked through the ether via the tubes CD. As the ether evaporates the silver cap is cooled until it reaches the dew point, when dew deposits on it.

As soon as dew is detected T is read. Another reading is taken when the dew disappears as the apparatus is allowed to warm up. The mean of these two values gives the dew point.

(ii) **Empirical hygrometers: Wet-and-dry-bulb hygrometer.**

Wet-and-dry-bulb hygrometer. Some skill is necessary in handling the hygrometer described above. The wet-and-dry-bulb hygrometer however requires no skill for making observations and is used extensively in meteorological work.

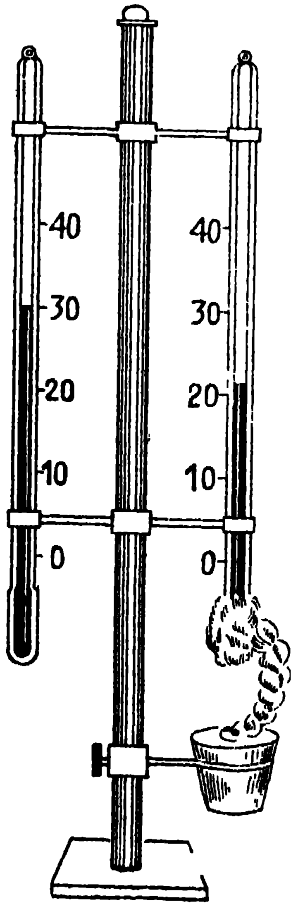


Fig. 119

It consists of two thermometers arranged on a stand, one of which is exposed to the air, while the bulb of the other is wrapped round with muslin or cotton wick that is kept moist by dipping into a small vessel of water (Fig. 119). The drier the air, the more rapidly will evaporation take place from the wet bulb and the lower will its temperature be. The thermometer with the dry bulb gives the temperature of the air. This temperature and the difference between it and that of the wet bulb enable us, by means of suitable tables, to obtain the percentage humidity, the dew point and the vapour pressure of air.

The instrument should be kept in a well ventilated place in a current of air.

Hair hygrometer. Human hair freed from fat and grease expands with increasing relative humidity. The difference between the lengths in dry air and in saturated air is of the order of 3%. One form of hair hygrometer is shown in fig. 120. The elongation of the hair is magnified by a suitable device. As the hair elongates, a pointer moves over a graduated scale. This hygrometer is a direct reading instrument, has an accuracy of about 3% and requires frequent checking.

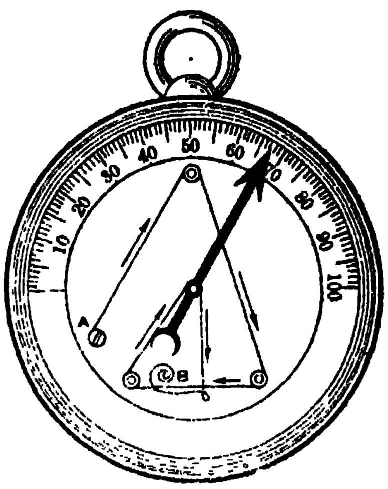


Fig. 120

133. Condensation of Water Vapour

The water that evaporates from the surface of the earth is returned to it in various forms. Cloud, mist, dew etc., are due to the condensation of water vapour in the atmosphere.

We have stated that when air containing water vapour is cooled gradually, it becomes saturated at a certain stage. When cooled further the air can no longer hold all the water vapour in it. Some vapour is condensed in the form of minute water drops. If, however, the air is pure it can contain much water vapour in excess of saturation. Under this condition the air or the moisture is said to be *super-saturated.*

For condensation of vapour to take place it is generally necessary that particles of some kind should be present in the atmosphere to act as centres or nuclei for the deposition of water molecules. Particles of dust may serve this purpose, but hygroscopic particles (such as common salt containing magnesium chloride as impurity) are very effective. The air near a town contains SO_2 molecules which come from burning coal. They are also hygroscopic and contribute largely to the formation of mist or fog near a town. Besides these particles moisture condenses easily on electrically charged molecules.

Let us consider some natural phenomena due to condensation of moisture in the air.

Dew. During the day objects on the surface of the earth are heated by direct radiation from the sun, and the air in contact with them is also heated. During the night the bodies lose heat by radiation and those which radiate well quickly cool below the temperature of the surrounding air. Air in contact with them cools too. If in this way the air cools below the dew point, a portion of the vapour is deposited as dew on the surfaces of adjacent bodies.

It is evident that the following conditions are necessary for the deposition of dew :

(*i*) *A clear sky.* If there is cloud in the sky at night, cooling of bodies by radiation is impeded.

(*ii*) *Absence of wind.* If there is wind, air near a cold object will not remain in contact with it long enough to cool below the dew point.

(*iii*) *Presence of copious moisture in the air.* If the initial humidity is high it does not require much cooling to reach dew point.

(*iv*) *The presence of objects which are good radiators, but bad conductors of heat.* Such objects cool rapidly and bring the air below its dew point. They should however be close to the earth as otherwise the cooled air will sink downwards and be replaced by warmer air from above.

Dew deposited on grass and leaves of plants is formed by water vapour given out by the leaves when the air around them has been cooled to the dew point.

Hoar-frost. When the dew point is below 0°*C* and the temperature of bodies lower still, water vapour in the air condenses directly as solid without passing through the liquid state. The solid is deposited on grass, etc., and is known as *hoar-frost.*

Fog and Mist. In a windless night the air at regions near the surface of the earth may be cooled below the dew point, when condensation of moisture occurs throughout the mass of air. The result is a *fog* or *mist.* There is no fundamental difference between the two. A mist in which one can no longer see objects at a distance of one kilometre, is usually spoken of as a fog.

The cooling is ordinarily due to radiation from the earth's surface. Hygroscopic particles play an important role in the formation of fog and mist.

Mountain mists or fogs are due to the contact of the cold air from the mountain sides with the warm saturated air of the valleys.

Clouds. When warm, moist air rises from a large surface of water, it gradually cools as it rises higher into the atmosphere. When the temperature of the air as a whole is reduced below the dew point, a cloud is formed by the condensation of vapour in the form of small droplets. The nuclei of condensation are minute salt particles originating from the sea and present in the air in sufficient numbers.

A cloud is thus a fog or a mist formed high up in the atmosphere. In the case of a fog or a mist the masses of air involved are often at rest or in very slow motion, but not so in a cloud.

The cooling which leads to the formation of a cloud may occur in a variety of ways. The chief causes are (i) the mixing of warm saturated air with a current of cold air and (ii) cooling by expansion as the air in its ascent moves to regions of lower pressure and increases in volume.

Experimet 37. The formation of cloud due to the cooling effect of expansion of air can be shown by the following experiment. A bell jar as in fig. 121 is connected to an exhaust pump through a tube passing through its air-tight stopper. A thermometer is also inserted. Partial vacuum is created within the jar by working the pump. When the inside and the outside temperatures become the same, some air is admitted from outside. The air expands suddenly and a fall in temperature will be noticed. If the admitted air is nearly saturated and contains dust particles, a cloud will be formed as soon as the air enters the jar.

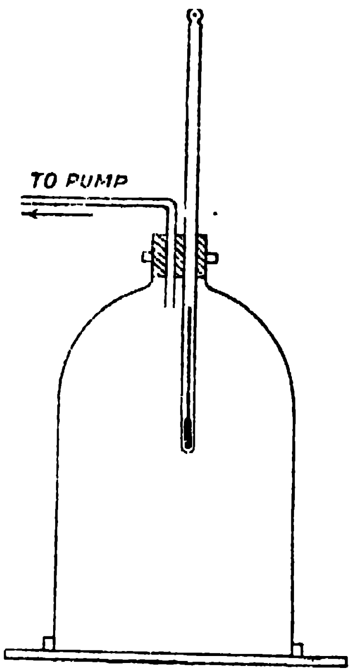


Fig. 121

Rain. The drops of water in a cloud coalesce and increase in size. Whether they will fall to the ground as rain depends on the condition of the air below the cloud, the upward speed of the cloud and various other factors.

The speed with which a rain drop moves towards the earth depends on its size. It is not a case of free fall under gravity, as the resistance due to air increases with velocity. A drop actually moves with a limiting speed, which depends on its size. Rain drops cannot be larger than 5·5 mm in diameter. Larger drops break up into smaller ones as they fall. Drops of this size fall with a speed of 8 metres per second or about 20 miles per hour. Other drops fall more slowly.

In order that a rain drop may fall to the ground, (i) the upward velocity of the cloud must be lower than the limiting velocity of fall of the drop, (ii) it must not encounter an upward moving mass of air of higher speed and (iii) must survive complete evaporation during the fall.

Water drops in a cloud are constantly going through the process of forming into large drops, breaking up into smaller ones and coalescing once again. Every time a drop breaks there is a separation of electricity. This is perhaps the principal source of electricity in a thunderstorm. Consequently thunderstorms are associated with heavy rainfall.

Exercises

✓1. Define *dew point* and *relative humidity*. How is the latter expressed ? What [illegible] Is the [illegible] the [illegible] that [illegible] ?

✓2. Describe a simple apparatus for measuring dew point. How would you determine relative humidity with its help ?

✓3. How is a wet-and-dry-bulb hygrometer used ? A weather report reads as follows :

"The maximum humidity was 95% and the minimum 36%. The maximum wet-bulb temperature was 28°*C*. The maximum temperature was 33°*C*."

Briefly explain the meaning of the report.

4. Explain how dew and fog are formed.

5. Answer the following :

(a) The bulb of a thermometer is wrapped round with cotton, which is wetted in turn with (*i*) water, (*ii*) ether, (*iii*) oil. How will the readings differ and why ?

(b) Explain why on a hot summer day, immediately after a rain, a block of ice on a cart appears to *steam* copiously.

(c) What becomes of the steam which a boiling kettle discharges into a room ?

(d) A piece of glass is dimmed when you blow on it with your mouth on a winter morning, but not on a summer noon. Explain.

CHAPTER 18

TRANSFERENCE OF HEAT

134. The Three Ways of Transference of Heat

There are three different ways in which heat can be transferred from one place to another, viz.,

(1) *Conduction.* When one end of an iron rod is put in the fire and the other end held in the hand, it will be noticed that this end gradually becomes warmer. Heat has passed from particle to particle along the rod from the warmer to the cooler end without any sensible movement of the particles. This process of transference of heat is called **conduction.**

(2) *Convection.* When the heated rod is removed from the fire and the hand placed a few inches *above* the hot end, it feels warm. Air heated by contact with the iron conveys the heat to the hand. Here the heat is actually carried by the particles of air, the movement of which is due to a change in density. Transference of heat due to the bodily motion of the warmer portions of a medium is called **convection.** It should be clear that convection is impossible in solids, but occurs in liquids and gases.

(3) *Radiation.* If instead of holding the hand above the hot end of the rod, it is held a few inches *below* the rod, the hand feels hot, but not to the same extent as before. Heat could not have been transferred by conduction or convection. The process by which heat is transmitted under this condition is called **radiation.** The heat that reaches the earth from the sun is transferred by radiation. We may say that radiation is the process by which heat passes from one body to another across a space without heating the space between the two

135. Conduction of Heat

Transference of heat by conduction requires a material medium. The particles of the medium take up heat energy from their neighbours on one side and hand it over to their neighbours on the other side. None of them leave their position.

Substances differ in their ability to conduct heat. One end of a glass rod may be put in the fire and brought to red heat while the other end is held in the hand. But if the glass is replaced by iron it may not be possible to hold it for long. This shows that heat is more easily transferred by conduction through iron than glass. We say that conductivity of iron is higher than that of glass.

Comparing conductivities of different solids—Ingenhausz's experiment. Conductivities of solids may be compared by Ingenhausz's method. The experimental materials (copper, iron, brass, glass etc.) are taken in the form of long thin rods of equal length, identical in area of cross-section and in surface finish. They are coated with wax and arranged as in fig. 122 with one end protruding into a tank in which water is maintained at the boiling point. Heat is conducted from the water through the rods and melts the wax. After a steady state has been reached, the wax is observed to have melted to different distances along the rods. Conductivity is proportional to the square of the distance over which the wax melts.

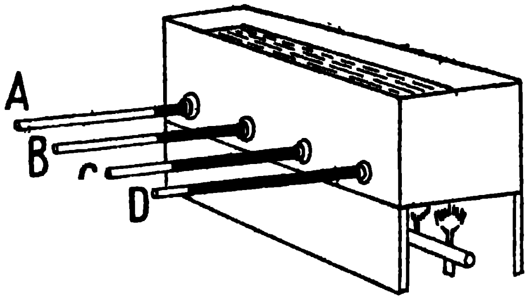


Fig. 122

Metals are the best conductors, silver heading the list with copper as a close second. Solids are in general better conductors than liquids, and liquids better than gases.

The low conductivity of water may be demonstrated in the following way. Load a piece of ice so that it sinks in water. Put it in a test-tube about two-thirds full of water (Fig. 123). Heat the tube near the top. The water at the top will be seen to boil before much of the ice has melted.

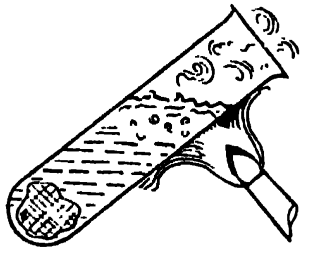


Fig. 123

Conductivity of gases is very low. Copper conducts heat 700 times better than water, and water, 25 times better than air. Wool, cotton-wool, felt etc. are bad conductors of heat and are used as thermal insulators. *They owe their low conductivity to the innumerable air pockets that they enclose.*

Some consequences of thermal conduction. (*i*) When we alternately touch a piece of iron and a piece of stone, both lying in the sun, the iron feels warmer than the stone though their temperatures are the same. This is due to the better conductivity of iron. Heat flows more quickly from it and produces an impression of greater warmth. When their temperatures are lower than that of our body, the iron gives the impression of being cooler due to the quickness with which heat is transferred from the body to it. If a piece of metal in arctic cold be touched with a finger, the finger freezes to the metal due to the large and quick transference of heat.

(*ii*) The 'warmth' of woollen garments is due to the innumerable air pockets they enclose. Because of the very low thermal conductivity of air, the garment as a whole has a low conductivity and retards the flow of heat through it. Two garments worn together will be more effective in preventing heat transference than only one of the same total thickness. The layer of air between the two further resists the flow of heat.

(*iii*) The temperature at which paper catches fire is much higher than that of boiling water. This makes it possible to boil a small quantity of water in a thin paper bag. Much of the heat supplied to the paper passes through it and while this heat raises the water to its boiling point, the temperature on the lower side of the paper remains below that at which the paper ignites.

(*iv*) When a clean copper gauze is lowered on the flame of a bunsen burner, it will be found that the flame does not get through the gauge (Fig. 124). The gauze conducts the heat away from the flame so rapidly that the temperature on the other side of the gauze is lowered below that at which the gas burns. Combustion is therefore confined to the part of the flame below the gauze. *When the gauze itself is heated to the ignition temperature, it ceases to be effective in preventing the combustion of the gas above it.*

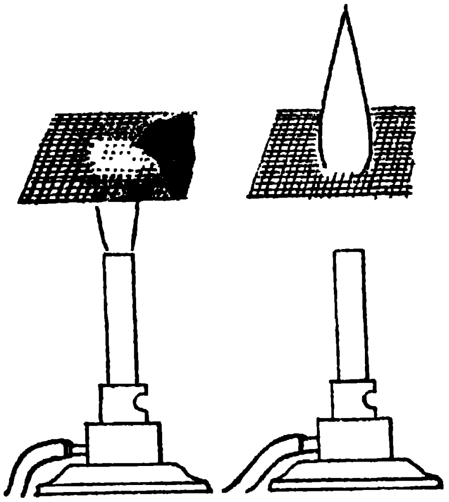


Fig. 124

When the gauze is held about an inch above a Bunsen burner and the flame lighted from top, it burns above the gauze (Fig. 124). The conductivity of the gauze prevents the flame from extending below the gauze.

(*v*) Asbestos, a fibrous mineral, is a bad conductor and is non-inflammable. It is extensively used for lining the walls of cookers and refrigerators. Boilers and steam pipes are often covered with an asbestos cement (asbestos fibre and plaster) to reduce leakage of heat.

(*vi*) Hard chalk deposits are found on the inside of boilers and kettles in areas where hard water is used. The deposit is a bad conductor. Unless it is removed from time to time much wastage of fuel is inevitable.

(*vii*) *Davy's Safety Lamp.* The principle illustrated in the flame-and-gauze experiment (Fig. 124) was utilized by Sir Humphrey Davy for the construction of a safety lamp for coal miners.

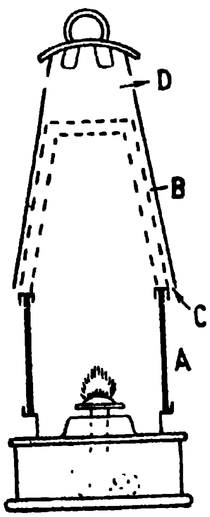


Fig. 125

It can be used in mines without risk of explosion even when combustible gases are present. It is an oil lamp with a wire gauze *B* (Fig. 125) arranged above the flame. The flame itself is surrounded by a glass cylinder *A* to increase the illuminating power of the lamp. Air enters the lamp at *C* and burnt gases leave the lamp at *D*.

In an explosive atmosphere the gases which penetrate to the flame are ignited and burn inside the gauze with a peculiar flickering. But the flame produced is unable to extend outwards through the gauze. The gauze conducts the heat away from the burnt gases so quickly that the explosive gas outside the lamp never becomes hot enough to explode.

Now-a-days miners use electric lamps. But the leader carries a Davy lamp to detect the presence of 'fire damp' (an explosive gas). If present in strong concentration, the gas burns with a blue flame.

(*viii*) The high thermal conductivity of copper makes it a material of choice in the construction of cooking utensils or small boilers, but its cost is relatively high. Aluminium being lighter and a better thermal conductor than iron is preferred for the construction of piston heads and engine cylinders.

(*ix*) An ice box or a refrigerator is constructed with a double wall. The space inside may be filled with saw dust or cork powder. Ice kept in saw dust melts slowly because of the bad conductivity of the innumerable air pockets enclosed between particles of saw dust.

136. Convection

Convection is possible only in liquids and gases, but not in solids. When a fluid is heated it becomes lighter and ascends, while its place is occupied by cooler portions of the fluid. The current so set up within the fluid is called **convection current.**

The ascending hot current carries the heat along with it. This process of transference of heat is called **convection.** It differs from conduction, since in the latter the medium through which heat flows is not displaced.

Simple experiments on convection. (*i*) Drop a tiny crystal of potassium permanganate into a beaker of water which is being heated by a burner from below. The water will be seen to move from the bottom up the middle and down the sides (Fig. 126). This shows that the water particles themselves carry heat from the bottom to the top while cooler particles come down to be heated and then move upward again.

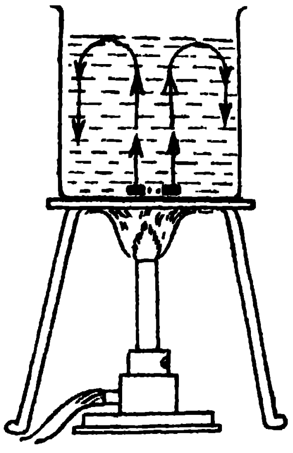


Fig. 126

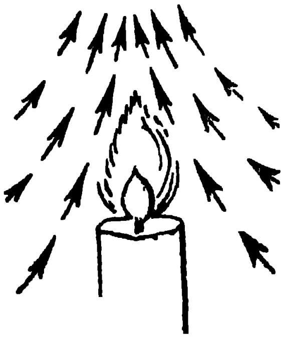


Fig. 127

(*ii*) When a candle burns a convection current is created in the air around the flame (Fig. 127). The air in contact with the flame is heated and ascends, while the surrounding cold air moves up to the flame. Thus the flame gets a continuous supply of oxygen.

(*iii*) In hurricane-lamps it is also the convection current which feeds the flame continuously with oxygen. Fig. 128 shows the path of the current. The burnt gases leave at the top.

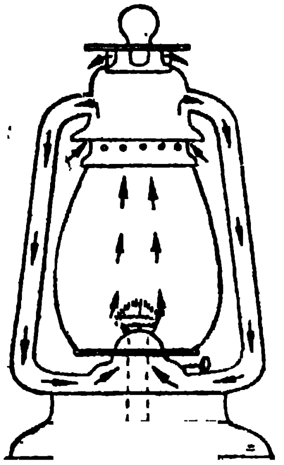


Fig. 128

(*iv*) The water cooling the engine of a motor car is kept in circulation by convection. As the water flowing in pipes surrounding the engine is heated, it rises to a tank at the front of the bonnet. Before returning to the engine it flows through small pipes cooled by air from a fan driven by the engine.

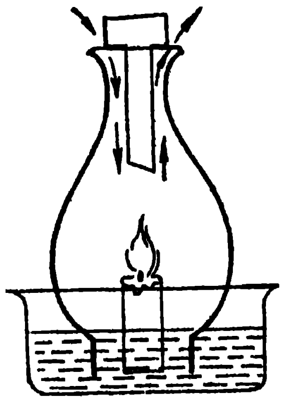


Fig. 129

(*v*) When a burning candle is placed inside a chimney in a way that does not allow air to enter through the bottom (Fig. 129) the flame soon goes out owing to lack of oxygen. If a T-shaped cardboard or metal sheet is placed as shown in the figure, the *candle continues to burn. The insertion of the sheet makes possible the flow of convection currents.*

(vi) Ocean currents are convection currents due to unequal heating of the water masses in the oceans by heat from the sun. Winds are also huge convection currents.

137. Radiation

In radiation heat passes from one body to another without the help of any material medium. The vast distance of 93 million miles from the sun to the earth is a void. (The density of matter in this region does not exceed one or two molecules per c.c.). Heat cannot therefore come from the sun to the earth by conduction or convection. It reaches the earth by radiation. In reaching the earth, heat from the sun passes through the earth's atmosphere ; but this does not heat the air. Maxwell defined radiation as the transfer of heat from a hot body to a cooler body without appreciable heating of the intervening space.

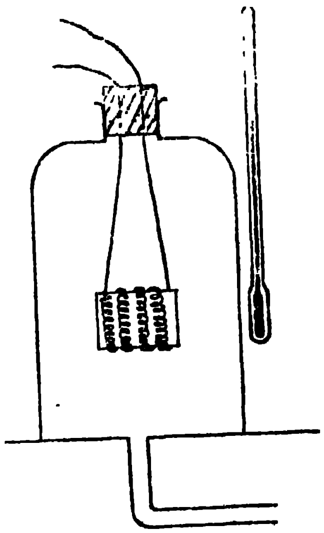


Fig. 130

Experiment to show transference of heat by radiation : A heating coil is suspended inside a bell jar through its lid (Fig. 130). Air from inside the jar can be pumped out by an exhaust pump. A thermometer is placed outside the jar. As the pump works a vacuum is created within the jar. The thermometer still shows a rise in temperature. There is no medium in the bell jar and as such the transference of heat by conduction or convection cannot take place. So there must be a third process by which heat is transferred. This process is radiation.

The form heat energy assumes when it is transferred from one body to another by radiation is called **radiant heat, radiant energy** or sometimes simply **radiation.** We now know that it is of the same nature as light and has all properties of the latter except that it does not produce the sensation of sight. Light and radiant heat travel with the same speed, i.e., 1,86,000 miles per second. During a total solar eclipse, darkness is accompanied by a fall in temperature.

Absorption and emission of radiation. The amount of heat lost from a body by radiation depends on

(*a*) its temperature and that of the surroundings, being proportional to the difference when it is not large ;

(*b*) the nature of its surface ;
(*c*) the area of the surface ; and
(*d*) the time.

When radiant heat falls on a body a part of it, depending on the nature of the material and the surface, is absorbed.

It has been found that **good absorbers of radiation are also good emitters.** Black is a better emitter or absorber than white. There is thus good reason for wearing white clothes in the tropics. A rough surface absorbs or emits radiation more than a polished surface. Calorimeters, therefore, have polished surfaces so that loss or gain of heat by radiation may be a minimum.

Prevost's Theory of Exchanges. It will be wrong to assume that a body does not radiate unless it is heated. When a slab of ice is suspended in a vacuum so as to reduce to a minimum the gain of heat by conduction and convection, it receives heat by radiation from the surroundings (if at a higher temperature than that of ice) and melts. If dry-ice (solid carbon dioxide) is similarly kept surrounded by ice, it evaporates. It cannot be logical to assume that ice radiates in the latter case, but not in the former. It is more reasonable to think that radiant energy is given out by ice in both cases. Melting takes place in the first case because the ice receives more heat by radiation than it gives out. In the second case, it gives out more heat by radiation to the dry-ice than it receives from the latter. In fact, "all bodies radiate heat and an exchange of radiation takes place between bodies at different temperatures. When a body emits more radiation than it receives from its surroundings, it cools. When a body receives more radiation than it emits, it warms up. When the radiation emitted is equal to the radiation absorbed, the temperature of the body remains constant. Even when equilibrium of temperature has been reached an exchange of radiation continues". The above idea of exchange of radiation between bodies at the same or at different temperatures is due to Prevost (1751-1839), professor of physics at Geneva. It is known as *Prevost's theory of exchanges*.

The vacuum flask. A vacuum or thermos flask (Fig. 131) consists of a double-walled vessel made of glass or metal. The air in the space between the walls is pumped out and the *surfaces bounding the vacuum* are silvered like a mirror. Heat cannot get across the empty space by conduction or convection since both of these require a material medium.

If there is a hot liquid in the flask, it loses little heat by conduction or convection. Heat loss by radiation from the inner wall is minimized by the silvering. The silvered outer wall being a poor absorber and a good reflector, returns the heat radiated by the inner wall back to the liquid. The same consideration apply to the entry of heat from the surroundings when the flask contains a cold liquid. Thus a liquid in the flask keeps warm or cold for a long time.

It should be noted that conduction and radiation cannot be entirely eliminated. A slow passage of heat due to these causes cannot be prevented. Hence the temperature of the liquid approaches that of the surroundings very slowly.

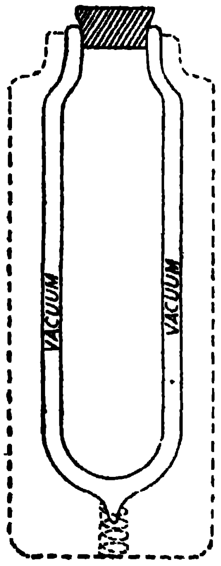


Fig. 131

Exercises

1. Illustrate with suitable examples the different methods of transference of heat. Bring out clearly the difference between them.

2. How would you show that different substances have different conductivities ?

3. Draw a sketch of Davy's safety lamp. Explain the use of the wire gauze surrounding the flame.

4. Give three examples of convection currents where they are made to serve some useful purpose.

Distinguish between conduction and convection.

5. Draw a neat sketch of a thermos flask, labelling clearly the silvered walls. Explain how heat transference is minimized in it when there is ice inside.

6. Hot water is placed in two identical jugs, one with a polished white surface, and the other with a black surface. Which one will cool more quickly and why ?

7. Clouds prevent loss of heat by radiation from the earth's surface. How does the presence of night clouds affect formation of mist or fog?

8. Answer why:

(*i*) The squirrel wraps its bushy tail round its body during its winter sleep.

(*ii*) Birds puff their feathers out on a cold day.

(*iii*) A hen, wishing to sit on eggs to hatch out chicks, casts feathers from her breast.

(*iv*) A house with a straw roof keeps cool in summer and warm in winter.

HEAT AND WORK

138. Heat is a Form of Energy

To the question 'What is heat' we may answer by saying that 'Heat is a form of energy ; it is not matter'. A body weighs the same whether it is hot or cold. If heat were a form of matter, the body would have weighed more when hot.

It is more difficult to demonstrate that heat is a form of energy. We have, however, been led to this conclusion by a variety of phenomena. When we hammer a piece of metal or bend a wire back and forth, it gets hot. Pressing a piece of steel on a grinder produces a crop of sparks. Examples of this kind where heat is produced by motion can be multiplied without limit. Steam engines (and other engines too) produce motion from heat. All these show that kinetic energy and heat are intimately related and one may be converted into the other. Heat must, therefore, be a form of energy. We now associate heat with the kinetic energy of motion of molecules. As the temperature rises, the kinetic energy of a molecule—whether in a solid, liquid or gas—also increases. Radiant heat, however, is a form of energy which exists apart from matter.

139. The Mechanical Equivalent of Heat

When we rub our hands together heat is produced; friction always produces heat. This shows that mechanical work can be converted into heat. The equivalence between heat and mechanical work was established by Joule whose experiments show that (*i*) **work can be fully converted into heat** and (*ii*) **whenever mechanical work is fully converted into heat, the heat generated is directly proportional to the work done.** The equivalence is independent of the way in which the work is derived or the means by which the transformation from work to heat is brought about.

If W units of mechanical work, on *full* conversion into heat, produce H units of heat energy, then

$$\frac{W}{H}=\textbf{constant}=J$$

where the constant J is called the **mechanical equivalent of heat** or **Joule's equivalent.** The numerical value of J however depends on the units in which W and H are expressed. If W is in ergs and H is in calories,

$$J = 4{\cdot}2 \times 10^7 \text{ ergs per calorie.}$$

This means that when $4{\cdot}2 \times 10^7$ ergs of work are *fully* converted into heat, the heat developed is one calorie, which can raise the temperature of 1 gm of water 1°C.

If W is in joules and H in calories, we have

$$J = 4{\cdot}2 \text{ joules per calorie.}$$

If W is in foot-pounds and H in British Thermal Units

$$J = 778 \text{ ft. lb per B.Th. U.}$$

If H is measured in lb.°C (centigrade heat unit)

$$J = 1400 \text{ ft.lb per lb.}^\circ C.$$

Thus if 1400 ft.lb of mechanical work, derived in any way, be *fully* converted into heat, it will raise the temperature of 1 lb of water 1°C.

140. Joule's Experiment for the Determination of J

The first exact determination of the mechanical equivalent of heat was made by Joule with the apparatus shown in Fig. 132. C is a cylindrical copper vessel containing water. Four radial partitions or vanes are fixed vertically inside the cylinder. A paddle can be rotated in the water about a vertical spindle (S). It carries eight sets of vanes which pass through spaces cut in the radial partitions as the paddle rotates. In the figure the radial partitions are represented by V_1 and the movable vanes by V_2. Fig. 133 presents a diagrammatic view of the cross section of the cylinder. To turn the spindle (S) it is attached by a pin (P) to a rod (R). Two threads wound round the rod (R) pass over two pulleys (P_1, P_2) and may be loaded at the other ends with known weights.

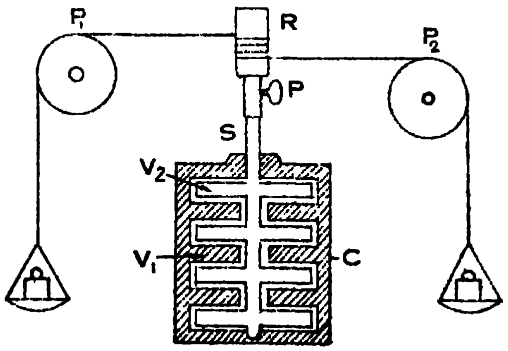


Fig. 132

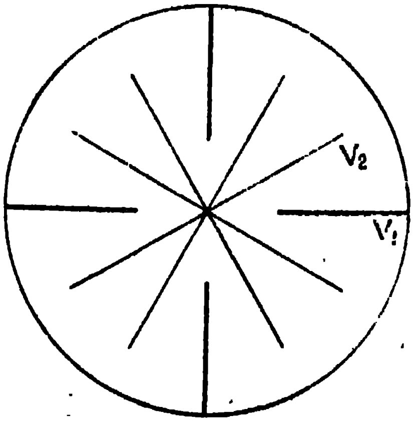


Fig. 133

When the weights are allowed to fall under gravity, the rod rotates and so does the paddle. As the water is churned its *kinetic energy is destroyed at the radial vanes* (V_1) and converted into heat. When the weights reach the ground, the pin (P) is detached and the weights are wound up again by turning a handle attached to the rod (R). As the churning continues water is gradually heated.

The elevated weights possess potential energy. When they reach the ground they have some kinetic energy. The difference of the two energies gives the kinetic energy imparted to the water. This is converted into heat.

Let the mass of each load attached to a string $= m$ gm,
distance through which each weight falls $= h$ cm,
acceleration due to gravity $= g$ cm/sec^2,
total number of times each weight is made to fall $= n$,
speed of a weight when it reaches the ground $= v$ cm/sec,
water-equivalent of the calorimeter with contents $= W$ gm,
initial temperature of the water $= t_1$ °C,
final temperature $= t_2$ °C.
Then, the potential energy of the weights $= 2mgh$,
their kinetic energy at the end of the fall $= 2 \times \frac{1}{2} mv^2$,
the mechanical energy converted into heat at each fall $= m\,(2gh - v^2)$,
the total energy converted into heat in n falls $= mn\,(2gh - v^2)$,
the total heat developed $= W\,(t_2 - t_1)$.

$$\therefore\ J = \frac{\text{mechanical work transformed}}{\text{heat generated}} = \frac{mn\,(2gh - v^2)}{W\,(t_2 - t_1)}$$

Corrections were made for the frictional losses and the loss due to radiation from the calorimeter. As in all calorimetric work, conduction and convection were minimized.

In one of Joule's experiments the weights were 30 lb each and fell 20 times in all, each time through an effective height of $5\frac{1}{3}$ ft. The water equivalent of the calorimeter and its contents was 14 lb and the rise in temperature, 0·59 °F.

$\therefore$ Heat generated $= 14 \times 0{\cdot}59 = 8{\cdot}26$ B.Th. U.
Work done $= 2 \times 20 \times 30$ lb. $\times\, 5\frac{1}{3}$ ft. $= 6400$ ft. lb,
whence $J = 6400$ ft. lb. $\div\, 8{\cdot}26$ B.Th. U.
$= 775{\cdot}7$ ft. lb/B. Th. U.

More accurate experiments to find the value of J were performed by Callendar, Rowland and others.

141. Heat Engines

We have considered the conversion of mechanical work into heat. The reverse transformation, i.e., the conversion of heat into mechanical work is also possible. When heat is converted into work the same equivalence holds ; for every calorie of heat converted into work we shall get nearly 4·2 joules.

Any device which converts heat into work is called a **heat engine.** The steam and the automobile engines, with which we are familiar in our every-day life, are examples of heat engines. But they are not the only ones of their kind. Gas turbines, jet-propelled engines and even the human body are examples of heat engines.

Principle of the heat engine. Heat engines, as we ordinarily know them, contain some substance, called the *working substance,* which takes in heat from a source at a relatively high temperature. From the state of high pressure at this high temperature it falls to a state of lower pressure, and in so doing does an amount of work, say, by pushing a piston. The heat equivalent of the work comes from the working substance itself and it is cooled to some extent. After having done the work the working substance gives out heat to another body, called the *sink,* which is at a lower temperature than the *source.* It again takes in heat from the source, does work, and hands over the balance of the heat to the sink. The process goes on in a cyclic order. We may say that in letting down heat from a higher to a lower temperature, a *part* of it may be converted into work under appropriate conditions.

If the heat absorbed from the source is Q_1 and that given out to the sink is Q_2, the difference Q_1-Q_2 has been converted into work. The magnitude of the work is

$$W=J\ (Q_1-Q_2).$$

The ratio $\frac{Q_1-Q_2}{Q_1}$ is called the **thermal efficiency** of the engine.

If the heat is absorbed at T_1 °K and given out at T_2 °K it may be shown that the efficiency cannot be greater than $(T_1-T_2)/T_1$.

142. The Steam Engine

The steam engine is one of the commonest and earliest forms of heat engines. Its action is as follows :

Action. The heat obtained by burning coal in a furnace vaporizes water in a boiler. The resulting steam is superheated and passed at a high temperature and pressure into a cylinder, where it forces a piston to move. It is in this process that heat is converted into work, and both the temperature and pressure of the steam diminish. The steam is then made to enter at the other end of the cylinder and pushes back the piston. Each time the piston is moved the steam does work and the exhaust steam passes out of the engine at a much lower temperature and pressure than that at which it entered. The to and fro motion of the piston is converted by a suitable device into the rotary motion of a shaft. This rotation is communicated by means of belting etc., to the machinery which the engine is required to drive.

The action may be understood by reference to Fig. 134 which represents a steam engine, greatly distorted in order to show how it works.

Description. High pressure, superheated steam enters the *slide valve chest* (SC) through the *steam pipe* (SP). Its further progress is regulated by the *slide valve* (V) which moves backwards and forwards over three openings (S_1, X, S_2) called *ports.* The outer ports [(S_1) and (S_2)] lead to the ends of the cylinder (C), while the central one, the *exhaust port* (X), goes round the back of the cylinder into the *exhaust pipe* (EP).

The *piston* (P) is within the cylinder. The *piston rod* (PR) is fixed firmly to it. The *valve rod* (VR) is fixed to the slide valve. At the places where the piston rod (PR) and the valve rod (VR) come out of the cylinder and the slide valve chest respectively, arrangements known as *stuffing boxes* (SB) are provided to prevent leakage of steam.

Action of the slide valve. When the slide valve is in the position shown in Fig. 134, steam enters the cylinder through the port S_1 and pushes the piston in front of it. The port S_2 and the exhaust port X are now in communication. Any exhaust steam which might be there on the other side of P passes through S_2 and X, and leaves the engine through the exhaust pipe EP.

The position of the piston controls the position of the valve (V). When the piston (P) has moved about one-third of its way down to the other end the valve (V) slides and closes one of the ports (S_1). No more steam can enter the cylinder (C) now. The steam which had already entered expands and does work in pushing the piston in front of it. It also cools in the process, the heat being converted into work. When the piston (P) moves to the extreme right the valve (V) slides further, opens the port to the right (S_2) and puts the other port (S_1) into communication with the exhaust port (X). Steam then enters through the port to the right (S_2) and the same process as above is repeated. The piston thus gets a to and fro motion.

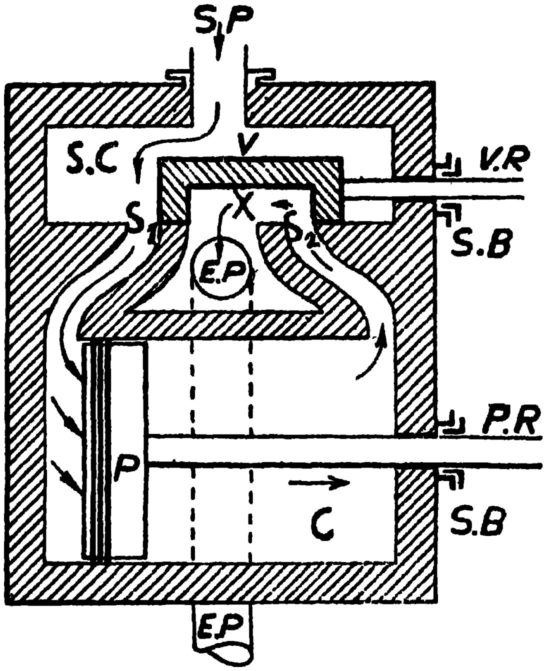


Fig. 134. SP=Steam pipe; SC=Slide valve chest; V=Slide valve; S_1,S_2=Steam ports; X=Exhaust port; EP=Exhaust pipe; C=Cylinder; P=Piston; PR=Piston rod; VR=Valve rod; SB=Stuffing box;

Efficiency. Only a part of the heat obtained by burning coal in the furnace is converted into work. A rough estimate of the relative heat losses in a steam engine is as follows :

Heat escaping through the chimney	10%
Heat lost by radiation	15% nearly
Heat carried by the exhaust steam	58%
Heat converted into work	17%

In spite of high pressure boilers and superheated steam the efficiency of a steam engine hardly exceeds 20%. Modern locomotive engines hardly reach an efficiency of 10%.

143. Internal Combustion Engines

In a steam engine the combustion of fuel takes place outside the cylinder; it may therefore be styled an external combustion engine. An **internal combustion engine** is one in which the combustion of the fuel is carried out in the engine cylinder. The fuel employed may be gaseous, such as lighting gas, power gas, petrol vapour etc., or a liquid, such as a fine spray of a combustible oil. According to the nature of the fuel they are sometimes called gas engines, petrol engines or oil engines.

The efficiency of an internal combustion engine is generally higher than that of a steam engine. In such an engine the heat produced by the combustion of the fuel in the cylinder is so great that the cylinder has to be cooled externally to carry away much of the heat.

144. The Petrol Engine

It is an internal combustion engine which burns petrol vapour inside the cylinder. Fig. 135 represents diagrammatically the essential parts of a petrol engine.

Description. Its cylinder (*Cy*) is open at one end. Beyond the other end of the cylinder there is a space equal to about one-fifth of the volume swept out by the piston. This is called the *clearance space*. In this space are located two valves (*I.V.* and *E.V.*) and the *spark plug* (*S.P.*). The first valve, called the *inlet valve* (*I.V.*), admits a mixture of air and petrol vapour into the cylinder. The other (*E.V.*) is the *exhaust valve* which lets out the spent gases after combustion. The spark plug provides an electric spark to 'fire' or explode the mixture

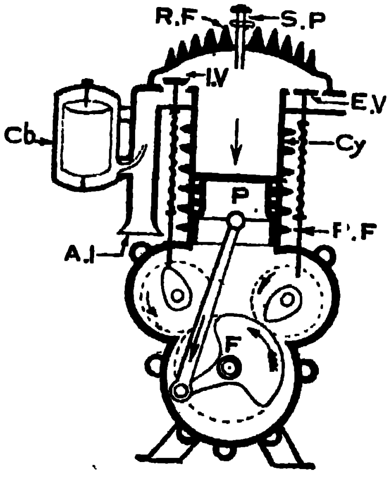


Fig. 135

Action. The action of the engine is started by setting the crank shaft into rotation by hand or some suitable starting device. The action of the engine consists of a successive series of changes which are repeated so long as the engine runs. *Six steps* may be distinguished in the cycle of changes, of which *four* are motions of the piston and are called *strokes*. Whereas in a steam engine each stroke is a working stroke, we shall see that of the four strokes in the cycle of changes in a petrol engine, only one is a working stroke. We now describe this cycle which is known as the Otto Cycle. It is represented digrammatically in Fig. 136.

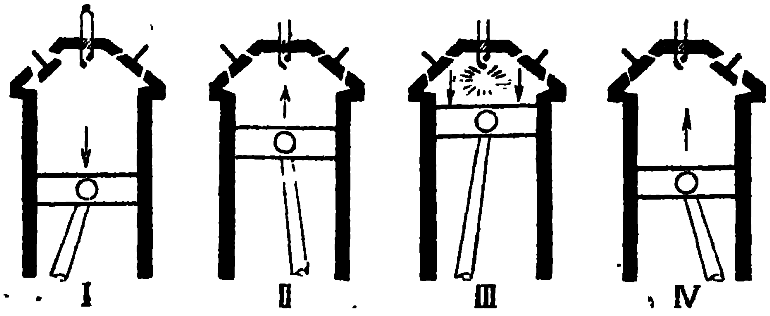


Fig. 136

(*i*) The *first step* is the first stroke or the **charging stroke** (briefly, '*suction*'). As the piston commences its outward motion the inlet valve opens and a combustible mixture of petrol vapour and air (the 'charge') from the carburettor is drawn into the cylinder. The exhaust valve is closed. The fly wheel supplies the energy necessary for the stroke (Fig. 136, I).

(*ii*) The *second step* is the second stroke or the **compressing stroke** (briefly, '*compression*'). Owing to the inertia of the fly-wheel the crank is carried round and the piston moves in. Both valves are closed. The piston compresses the charge into the clearance space (Fig. 136, II). The pressure and the temperature of the compressed mixture rise considerably. The fly-wheel supplies the energy necessary for compression.

(*iii*) The *third step* consists in the firing of the explosive mixture by an electric spark from the spark plug. The pressure and temperature rise to very high values. (This step is to be looked upon as the absorption of heat from the source by the working substance. The explosive combustion of petrol vapour liberates a large amount of heat in the cylinder).

(*iv*) The *fourth step* is the third stroke or the **working stroke** (also called the '*power stroke*'). The hot gases at high pressure force the piston outwards (Fig. 136, III). The gases do work and fall in temperature and pressure. Part of the heat they contain is converted into work.

(*v*) The *fifth step* is the opening of the exhaust valve after the piston has completed its outward motion during the working stroke. The spent gases, yet very hot, rush out of the cylinder through the exhaust valve. (This step may be looked upon as the letting out of heat by the working substance.)

(*vi*) The *sixth step* is the fourth stroke or the **exhausting stroke** (also called the '*scavenging stroke*'). The exhaust valve remains open throughout the stroke. As the piston moves in, it pushes the spent gases out through the exhaust valve. The fly-wheel supplies energy during this stroke. (Fig. 136, IV)

At the end of this stroke the exhaust valve closes, the inlet valve opens and the engine is ready to begin its next cycle again.

For such points as the mixing of fuel in the right proportions, the arrangement for the best timing of the spark and the opening and closing of the valves, the reader should consult a book on heat engines.

Efficiency of engines. The petrol engine of a motor car has an efficiency of about 25%. Diesel engines, a class of internal combustion engines which use liquid fuel and bring about combustion not by a spark but by high compression of the air, reach an efficiency of about 35%. In steam turbines, another class of heat engines, the efficiency is about 25%. Modern jet-propelled engines have an efficiency of about 10% ; but it is being improved rapidly.

A rough estimate of the heat expenses on different accounts of a petrol engine is as follows :

Heat lost from the cylinder by conduction, convection and radiation	36%
Heat passing out of the exhaust valve	35%
Heat loss due to friction	8%
Heat converted to work	21%

The human body is also a heat engine where food is the fuel and generates heat within the body by 'burning' (i.e. combining with oxygen).

Exercises

1. What reasons have we to think that heat is a form of energy ?

2 What do you mean by the expression that the mechanical equivalent of heat is 778 ft. lb. per B.Th. U. ?

Briefly relate how this quantity has been determined.

3. Explain with the help of a neat diagram the action of a steam engine.

4. Draw neat diagrams to illustrate the 'strokes' of a petrol engine, and briefly describe how the engine acts.

5. A mass of 10 gm moving with a speed of 300 metres per sec. is suddenly brought to rest. If the whole of its kinetic energy is converted into heat, find (*i*) the heat developed and (*ii*) the rise in temperature of the body assuming that the heat developed is fully retained by it. (Sp. heat $= {\cdot}03$).

(*Ans* : 107·1 calories ; 357°*C*.)

6. Water overflows a spillway of height 975 ft. Calculate the rise in temperature of the spilled water at the foot over that at the crest. $J=780$ ft. lb. per B. Th.U. (*Ans* : 1·25°*F*)

7. Earth's atmosphere retards the speed of a 42 kgm meteor from 18 kilometres per minute to 3 kilometres per minute. Calculate in calories the heat developed due to the change in speed. $J=4{\cdot}2\times10^7$ ergs/calorie. (*Ans* : 437500)

CHAPTER 20

LIGHT AND ITS RECTILINEAR PROPAGATION

145. Introductory

Light is the external physical agency which produces the sensation of sight. When light from an object enters the eye and reaches a screen, called the *retina,* placed to the rear of the eye, it excites the nerve endings with which the retina is copiously supplied. Due to this excitation electrical impulses are carried along the nerves to the brain and give rise to sensations of brightness and colour. Unless light from an object reaches the eye, the object cannot be seen. *Light itself is invisible.*

Some Definitions. A body which emits light, such as the sun, a lamp, a candle-flame etc., is a **self-luminous** body or a source of light. Most bodies are however **non-luminous.** They are made visible by the light which they receive from self-luminous bodies.

A substance through which light passes is called a **medium.** When the properties of the medium are the same at every point, it is called a **homogeneous** medium. If the properties vary, the medium is heterogeneous. A medium is said to be **transparent** when bodies can be seen distinctly through a sheet of the medium. A substance through which light cannot pass except when the thickness is very small, is called **opaque**. There are substances through which light can pass to a greater extent, but objects cannot be seen clearly. Such substances are called **translucent.** Ground glass, oiled paper etc. are examples of translucent substances.

The path along which light travels in going from one point to another is called a **ray.** In a homogeneous medium a ray may be represented by a straight line with an arrow-head indicating the direction in which the light travels. A group of rays taken together is called a **pencil of rays.** A group of parallel rays is called a **parallel pencil** [Fig. 137 (i)]. If the rays proceed to meet at a point, it is called a **convergent pencil** [Fig. 137 (ii)]. When the rays of a pencil move outward from a point they constitute a **divergent pencil** [Fig. 137 (iii)]. The point F of convergence or divergence of the pencil is called its *focus.*

When the dimensions of a source of light are very small compared with the distances involved it is called a **point source.** A star is a point source of light though it is millions of times bigger than the earth.

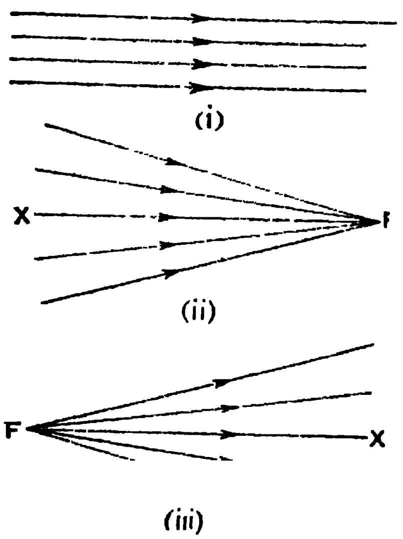


Fig. 137

To see any point of an object a pencil of rays from the point must reach the eye. When the object is self-luminous every point of it radiates light in all directions. Light travels outwards with a velocity of 186,000 miles per second, or in cgs units, 3×10^{10} cm per second. When light falls on a non-luminous object some of it is scattered in all directions. If any of this scattered light reaches the eye of an observer he sees the object.

146. Rectilinear Propagation of Light

In a homogeneous medium, such as air, light travels from one point to another in a straight line. This behaviour of light is known as rectilinear propagation. It can be simply demonstrated as follows.

(i) Fasten one end of a string to a hook in a fixed post and pull the string to make it taut. Look along the string. You will see the hook to which the string is fastened. If you look in any other direction you do not see the hook. This shows that the light from the hook reaches your eye along a straight line.

(ii) Take three or more cards with a small pin-hole in each of them (Fig. 138). Pass a thin metallic wire through the holes and keep the wire stretched. Clamp the cards in position at a distance from one another. Cut off the wire and remove it, taking care that the cards are in no way displaced. If the wire did not sag under the weight of the cards, clearly the pinholes would be in a straight line. If a candle-flame is placed in front of the hole in the first card an observer will see the light through the series of holes. If any of the cards is displaced in its own plane light can no longer be seen.

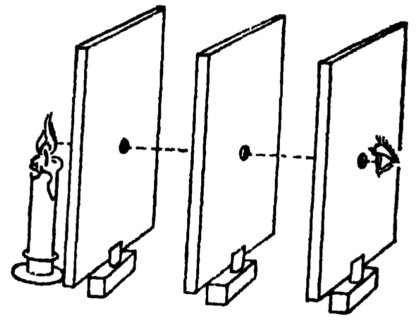


Fig. 138

The medium which extends between the source and the observer's eye is air. Over such a small region air is homogeneous. The experiment therefore shows that *in a homogeneous medium light travels in a straight line.*

(iii) When a small obstacle is placed between an observer's eye and a small distant source of light, the latter is cut off from view. This also shows that light travels in straight lines.

(iv) Most of you would have noticed the sun shining into a room through a small hole in a window. If dust particles are stirred up in the room they make the path of the sun-light visible. You will find that this path is straight. The path of the light can also be made visible by blowing smoke.

147. The Pin-hole Camera

An interesting application of the property of rectilinear propagation of light is the pin-hole camera. To understand its action make a small hole H (Fig. 139), about one millimetre in diameter, in an opaque screen S and place a luminous body, such as a candle-flame AB, on one side of it. On the other side place a ground glass screen G. A narrow pencil of rays starting from each point of the source will, after passing through H, form a small patch of light on G. An image $A'B'$ will, therefore, be formed on G. It will be inverted because the rays from different points on AB cross one another at H.

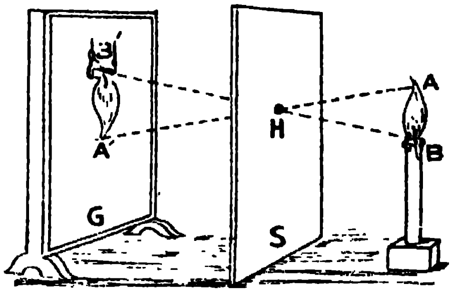


Fig. 139

If the hole H is gradually increased in size, the patch of light formed on G by the pencil of rays starting from any point of AB also grows larger. The image becomes more and more blurred due to the superposition of light from different points of the object. Distinctness of the image, therefore, diminishes with increase in size of the hole. Brightness of the image, however, increases with increase in size of the hole, as more light now reaches the screen G.

The above explains the action of the pin-hole camera. It consists of a light-tight box with a small hole at the centre of one of its walls. On the opposite wall is placed a photographic plate to receive the image. As the light passing through the hole is very small, a long exposure, of about half-an-hour, is required, No focusing is necessary as in the ordinary camera,

148. Circular or oval Patches of Sun-light beneath Trees

We have just seen how a small hole forms an inverted image of an object. The shape of the hole does not affect the image if the image is received on a screen at a reasonable distance from the hole. The circular or elliptic patches of light seen on the ground beneath a tree are images of the sun formed by the sun-light passing through the interstices of leaves. When the beam strikes the ground perpendicularly, the image is circular. When it falls obliquely on the ground, the image is elliptic. The shape of an interstice through which the light passes, does not affect the appearance of the image; but its size determines the brightness of the image.

Such patches are often seen in a room when the sun shines through a small hole in a door or a window.

149. Shadows

Another consequence of rectilinear propagation of light is the formation of shadows.

A. Point source of light. When an opaque body *AB* is placed in front of a point source *S* of light and a screen *PQ* put up at a distance behind the obstacle, a sharp shadow is cast on it. No point within *MN* receives any light at all (Fig. 140).

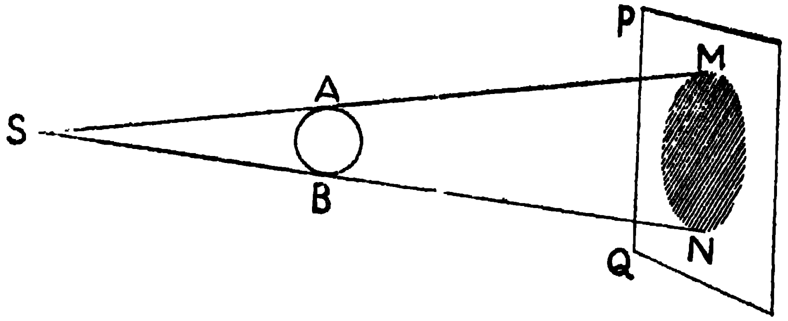


Fig. 140

B. Source of finite extension. When the source of light has a finite size, the shadow it casts of an opaque body, consists of a central region of perfect darkness, called the **umbra**, and a less dark outer region, the **penumbra**. No rays from any part of the source reaches the umbra. A point on the penumbra, however, gets light from some part of the source. Away from the umbra the illumination gradually increases till it reaches the full brightness outside the shadow. The shadow is never sharp.

(i) **Source smaller than the obstacle.** *SS'* is the source, *AB* the obstacle and *PQ* the screen (Fig. 141). The points between

M and *N* are in perfect darkness and constitute the umbra. The region inside *M′N′* but outside *MN* will be less dark. It is the penumbra. The penumbra shades into the complete darkness of the umbra on one side, and into complete brightness outside the shadow on the other.

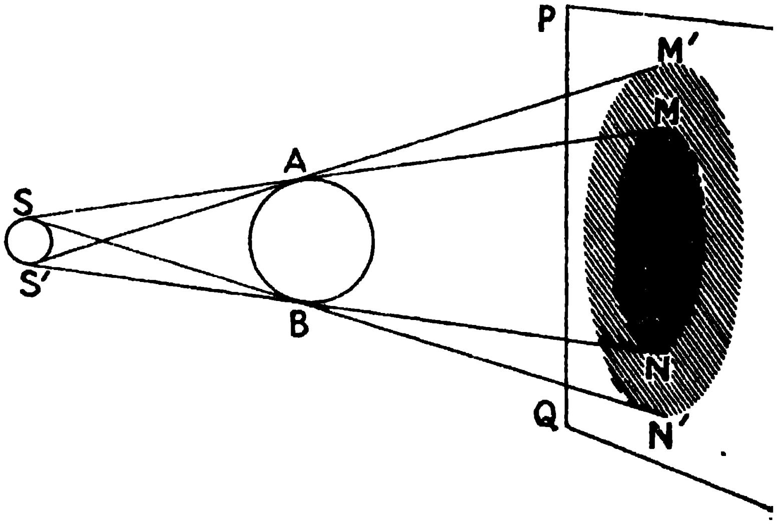


Fig. 141

(ii) **Source larger than the obstacle.** It is clear from fig. 142 that the umbra forms a *cone* tapering to a point behind the object. *ABF* is the *umbral cone*. The size of the umbra depends on the position of the screen. Beyond the point *F* the whole of the shadow consists only of the penumbra. Viewed from a point within the extension of the umbral cone the source looks as shown in fig. 142 (*b*), lower diagram. From a point in the penumbra the

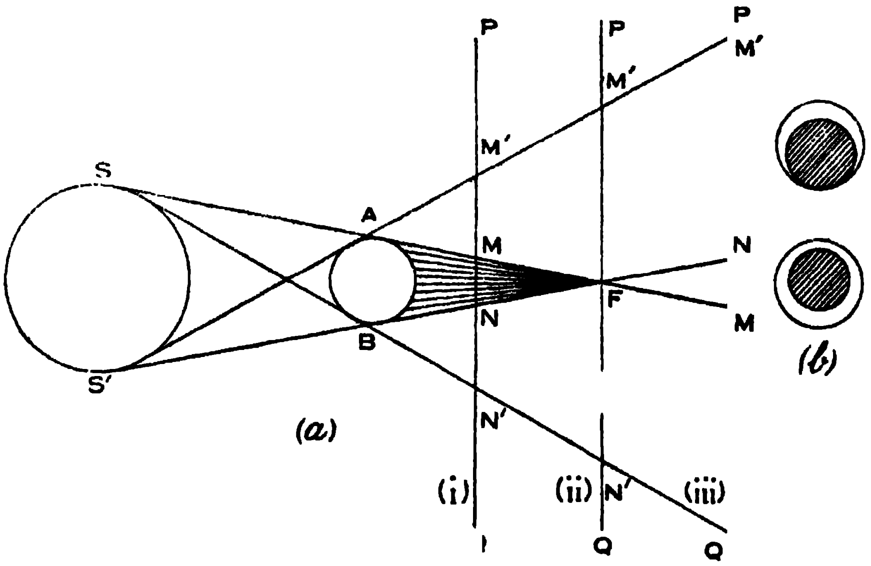


Fig. 142

source looks as in the upper diagram of fig. 142 (*b*); but the size of the crescent changes with position.

150. Eclipses

Eclipses are due to the shadow of one celestial body falling on another.

A. Lunar eclipse. The shadow of the earth cast in space by the sun consists of an umbral cone surrounded by the penumbra (Fig. 143). Fig. 144 shows the sun *S*, the earth *E*, the moon *M* and the umbral cone of the earth. p_1, p_2 are the limits of the penumbral region. u_1, u_2 are the limits of the umbral cone.

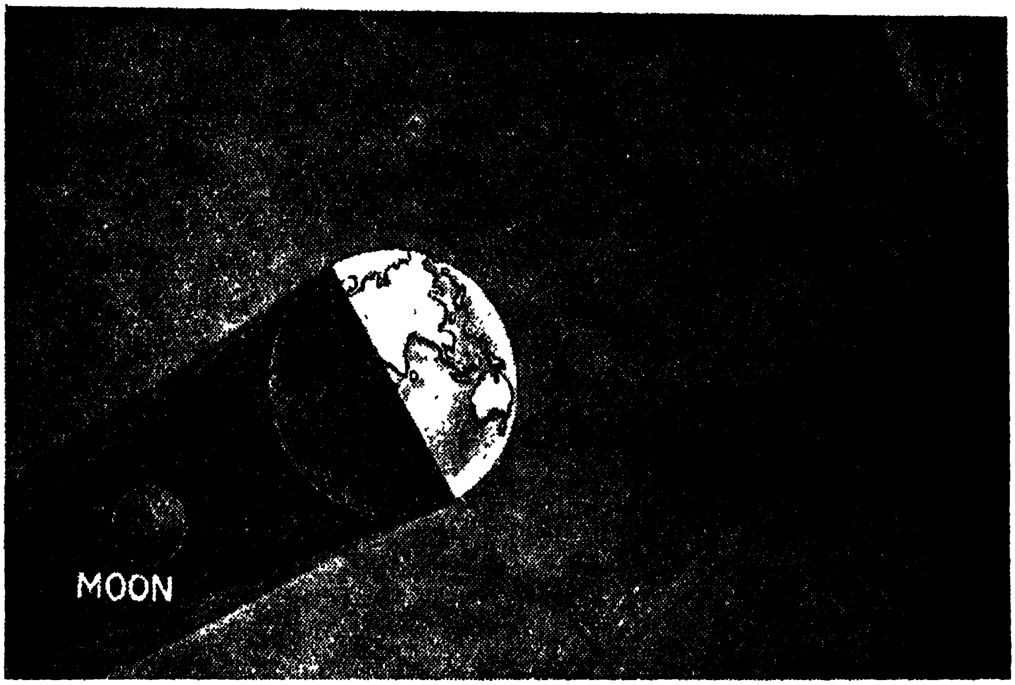


Fig. 143

(The sun in 93 million miles away from the earth and 109 earth diameters across. The moon is a quarter million miles away and one-fourth of the earth in diameter. The length of the umbral cone of the earth is about $8{\cdot}7 \times 10^5$ miles, and it is 6000 miles across at the distance of the moon.

The large differences in magnitude make it impossible to draw a diagram to scale within a small compass and yet to show the details clearly. Fig. 143 to 147 are, therefore, schematic.)

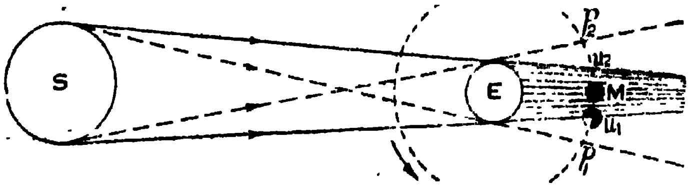


Fig. 144

If the moon during its motion round the earth enters the umbral cone an eclipse occurs. When the moon lies entirely within the umbral cone (*i.e.*, within u_1 u_2) the eclipse is *total.* When it is partially inside and partially outside this cone, eclipse is *partial.* As the moon moves from p_1 to u_1 the light it gets from the sun gradually diminishes. So the onset of an eclipse is preceded by a diminution in the brightness of the moon. But this very often goes unnoticed.

Lunar eclipses can occur only at full moon, when the earth is between the sun and the moon. The orbit of the moon is inclined at about 5° to the orbit of the earth. The path of the moon may therefore lie entirely outside the shadow cone of the earth and we may not get an eclipse at a full moon. There are two points at which the path of the moon cuts the plane of the earth's orbit. These points are called the *nodes.* The line joining the nodes rotates slowly. *A lunar eclipse takes place only when the nodes, the sun and the earth are in one line, while the moon is at or near the node farther from the sun.*

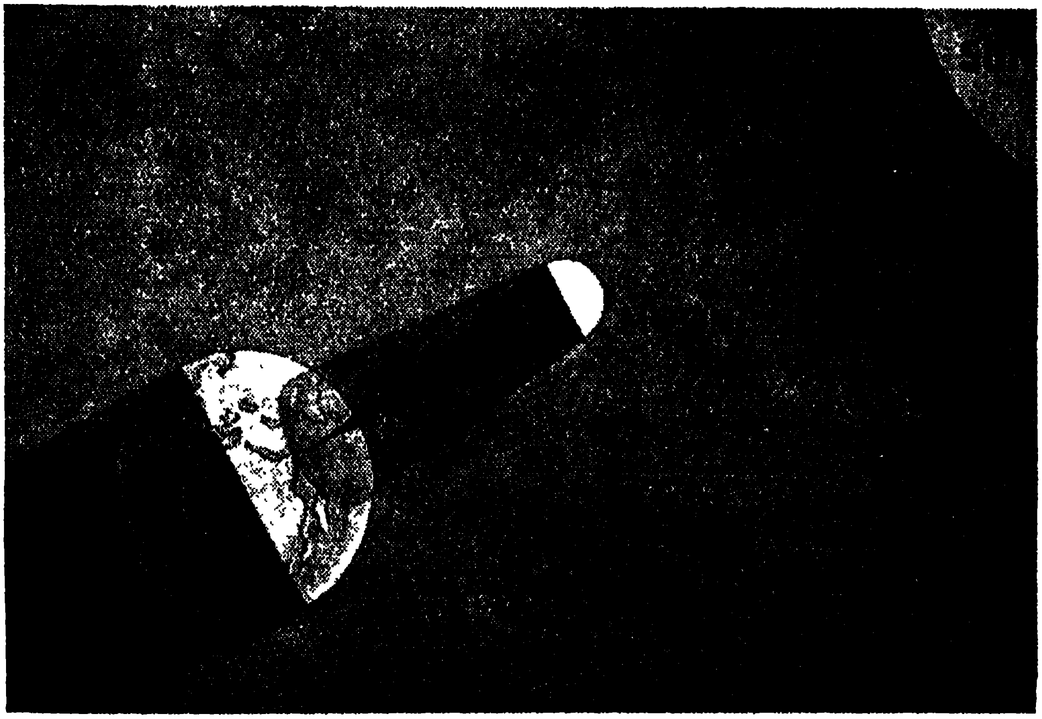


Fig. 145

B. Solar eclipse. The shadow of the moon cast in space by the sun consists of an umbral cone surrounded by a penumbra (Fig. 145 & 146). At a new moon, when the moon is between the sun and the earth, the shadow of the moon may fall on the earth and cover a part of it. From a place on the earth covered by the umbral cone of the moon, no part of the sun can be seen. At such places (area *T*; fig. 146) we have a *total* solar eclipse. As the earth is very near the apex of the moon's umbra, only a small area can have a total eclipse at any instant.

From places covered by the penumbra, (area *PP*, lightly shaded ; fig 146) only a part of the sun will be seen. Such places will, therefore, have a *partial* solar eclipse. An area of total eclipse is surrounded by a wider area of partial eclipse.

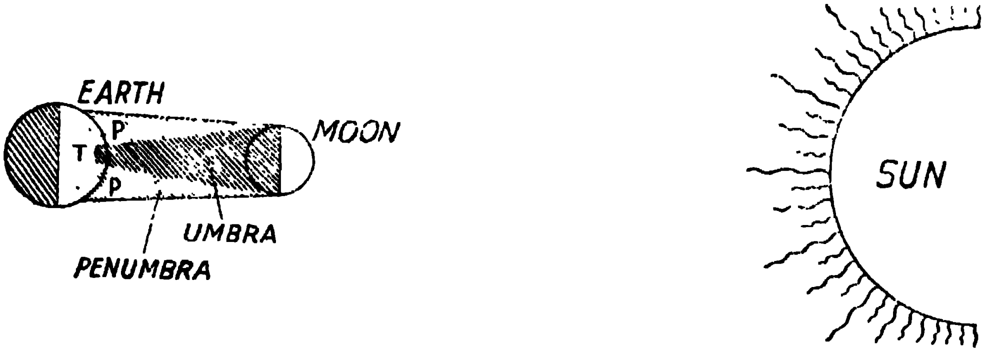


Fig. 146

A solar eclipse does not take place at each new moon. For though the moon passes between the sun and the earth at a new moon, it may be above or below the line joining the other two. Hence its shadow may not fall on the earth. *A solar eclipse takes place only when the nodes, the sun and the earth are in one line, while the moon is at or near the node closer to the sun.*

Annular solar eclipse. The path of the moon round the earth is an ellipse. The length of the umbral cone of the moon has a value intermediate between the maximum and minimum distances of the moon from the earth. If at a solar eclipse the earth is more distant from the moon than the apex *F* (Fig. 147) of the umbral cone the eclipse will be **annular** over the region of the earth's surface lying within the extension *A B* of the umbral cone. For any point within this region the moon cuts off light from a circular region of the solar disc, and the sun looks as in the lower diagram of fig. 142 (*b*). The eclipse is said to be *annular*. In the penumbral region *AA′* and *BB′* (Fig 147) of the earth the eclipse is *partial*.

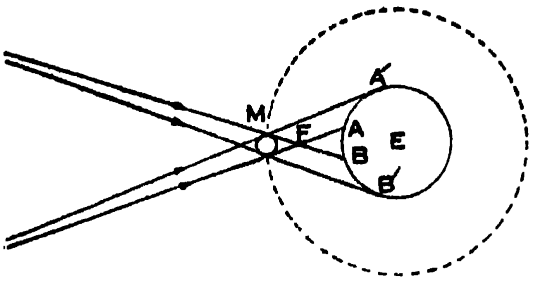


Fig. 147

To have an annular *lunar* eclipse the diameter of the earth's umbral cone should be smaller than the diameter of the moon. The moon has, therefore, to be much farther away from the earth than at present—in fact, over three times its present distance.

Exercises

Define a ray and a point source of light.

1. How would you demonstrate that light travels in straight lines in a homogeneous medium ?

2. Explain with a diagram the action of a pin-hole camera. What would be the effect if the size of the pin-hole is increased ?

3. What are umbra and penumbra ? Explain their formation when the source of light is larger than the obstacle.

4. What is an umbral cone ? In a dark room there is a hollow glass globe of $2''$ diameter containing an electric bulb. An opaque ball of $\frac{1}{2}''$ diameter is placed such that its centre is at a distance of 4 in. from the centre of the globe. Draw a diagram showing the umbra and penumbra and find out the length of the umbral cone. Why is not the penumbra uniformly bright all over.

5. Explain with diagrams how solar and lunar eclipses take place. What is an annular eclipse ? Why cannot it occur during a lunar eclipse ?

6. An aeroplane flying high up in the air may not cast a detectable shadow on the ground. Explain why ?

7. Light of the sun shining through the leaves of a tree form circular or oval patches on the ground. Expain how they are formed.

CHAPTER 21

PHOTOMETRY

151. Photometric Concepts

Light is a measurable quantity. Just as an electric current may be regarded as a flow of electric charge, *light may be regarded as a flow or flux of radiant energy.* **Photometry** is the science of measuring light. The concepts used in photometry may be understood with the help of the following simple illustrations.

When a candle in a room is replaced by an electric lamp the latter sends out a much larger quantity of light into the room. More light flows out of the electric lamp than the candle. In other words, the electric lamp produces a much greater flow or flux of light, i.e. **luminous flux,** than the candle.

When the electric lamp is the source, the amount of light falling on different objects in the room is greater than when the candle is the source. In other words, **illumination** is increased.

152. Photometric Definitions

The above concepts lead to the following definitions :

(i) **Luminous flux.** *The quantity of light energy which a source sends in unit time through a surface is called the luminous flux through the surface.*

(ii) **Luminous Intensity or Illuminating Power.** If all luminous sources emitted light equally in all directions the power of any source could be specified by the *total luminous flux* it emits (i.e., sends through a closed surface surrounding it). But this is not generally the case. It is therefore necessary to have some means of specifying the *rate* at which a source emits light in *any given direction.* The quantity *luminous intensity* is introduced to meet this necessity.

Luminous intensity or illuminating power of a source is defined as *the quantity of light energy which the source sends in unit time through unit area placed perpendicular to the rays at a distance of unity from the source.*

Units of luminous intensity and flux. The unit of luminous intensity (or illuminating power), as originally defined, is the **candle,** which is the *luminous intensity or illuminating power in a horizontal direction of a spermaceti candle, six to the pound,* *burning at the rate of* 120 *grains per hour.* This unit was long ago found to be very unsatisfactory. It has now been replaced by the **international standard candle,** which is *one-tenth of the intensity at right angles to the flame of a Vernon-Harcourt pentane lamp* burning under prescribed conditions of humidity and pressure. The lamp burns pentane vapour mixed with a certain definite amount of air. Under defined conditions it provides a very constant source of light. The *luminous intensity of any source of light measured in terms of this standard* is called the international candle power or simply **candle power** of the source.

The unit of luminous flux is the **lumen**. It is defined as the *luminous flux passing through unit area of a sphere of unit radius from a uniform point source of one candle power situated at the centre of the sphere.* As the total area of a unit sphere is 4π, *the total luminous flux from a uniform source of one candle power is* 4π *lumens.*

(iii) **Illumination** or **intensity of illumination** of a surface is defined as *the quantity of light energy falling normally per unit area per unit time.* To obtain its measure divide the total luminous flux falling normally on an area by the area.

The unit of illumination (or intensity of illumination) might have been *one lumen per square centimetre* (called the *phot*); but it is an impracticably large unit. The unit in use is the **lux** (or the **metre-candle**). It is the illumination produced on a surface at a distance of one metre from a point source of one candle power the surface being normal to the rays. Note that this is *one lumen per square metre.* When the surface is held at a distance of one foot from a point source of one candle power, the resulting illumination is one **foot-candle** or *one lumen per square foot.* The illumination of $OPQR$ in fig. 148 is 1 foot-candle.

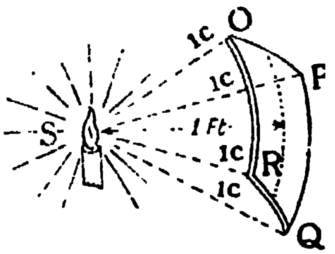


Fig. 148

1 lux = 1 lumen/sq. metre = 1 metre-candle

1 phot = 1 lumen/sq. cm. = 10^4 lux

1 foot-candle = 1 lumen/sq. ft. = 10·764 lux

(iv) **Brightness.** The idea of luminous intensity is applicable only when the source is so small that it may be treated as a point. For extended sources the corresponding quantity is called 'brightness'. The **brightness** of an extended source *in a given direction is defined as the amount of light energy emitted* the screen. The distances of one or both of the sources are then adjusted so that both the shadows appear equally bright. Since each shadow is illuminated by light from only one source, we have, when the shadows are equally bright, $I_1/r_1^2 = I_2/r_2^2$ where I_1 and I_2 are the intensities of the sources and r_1 and r_2 their distances from the screen respectively. The sources should be placed close to the normal from *R* on *S*, so that incidence of light on *S* may be practically normal. The experiment must be performed in a dark room.

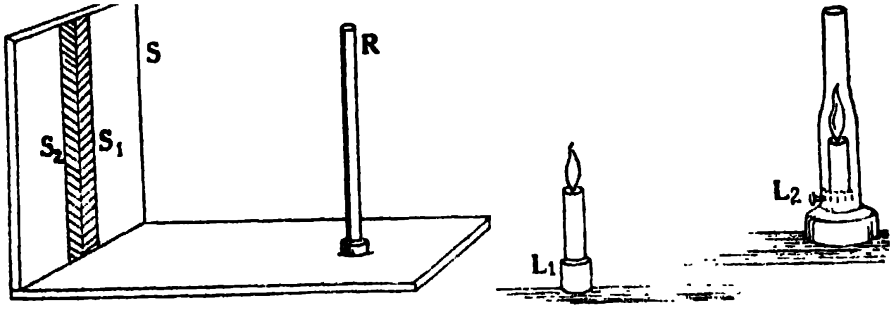


Fig. 151

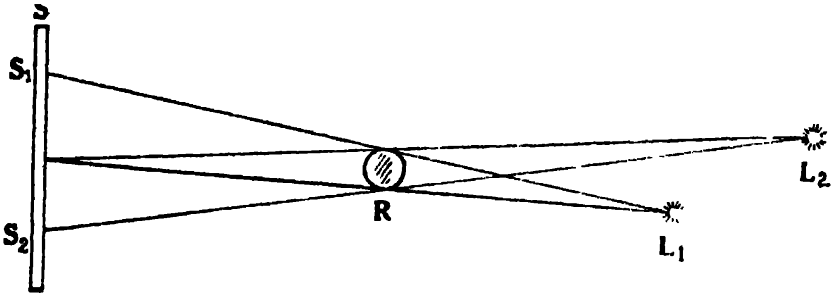


Fig. 152

156. Direct reading Photometers

The usual photometric methods make use of the eye as the instrument of comparison. Various electrical effects produced by light can also be used for the purpose. One such effect, known as the **photo-voltaic effect,** has led to the development of a direct reading instrument which indicates the illumination which a screen on it receives. The screen consists of a thin layer of cuprous oxide on a copper plate, covered with a thin layer of metal (Fig. 153). When light reaches the copper oxide film through the thin layer of metal, electrons are emitted from it and collect on the metal layer. The potential difference thus set up is proportional

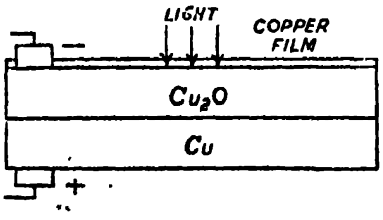


Fig. 153

to the intensity of the incident light (or the illumination on the surface). The potential difference is of the order of 1/20 volt and may be made to move the needle of a galvanometer on a scale calibrated in foot-candles. **Exposure meters** used in photography are based on this principle.

157. Experiments on Photometry

Experiment 38. *To compare the intensities of two sources of light.*

Apparatus : A Rumford shadow photometer, two sources and a scale.

Method : The experiment should be done in a dark room. Place the two sources in front of the photometer such that the shadows are formed side by side on its screen. Adjust the distances of the sources so that the shadows appear to have equal illumination.

Measure the distances of the screen from the sources. Let I_1 and I_2 be the intensities of the sources and r_1 and r_2, their respective distances from the screen. Then

$$I_1/I_2=r_1^2/r_2^2$$

For accuracy, the following precautions should be taken :

(i) The sources should be placed close to the normal drawn from the opaque obstacle to the screen so that the incidence of light may be practically normal.

(ii) For easy comparison the shadow due to one source should exactly touch but not overlap the other shadow.

(iii) The room must be absolutely dark.

Experiment 39. *To verify the law of inverse squares.*

Apparatus. A Bunsen grease-spot photometer, several lamps of equal candle power and a scale.

Description. A Bunsen photometer consists of a white paper screen S (Fig. 154) with a sharply defined grease spot P at the middle. When the screen is held near a window and observed from inside the room the spot appears brighter than the rest of the paper. If it is viewed from outside the spot appears darker. When the illumination on the two sides are the same

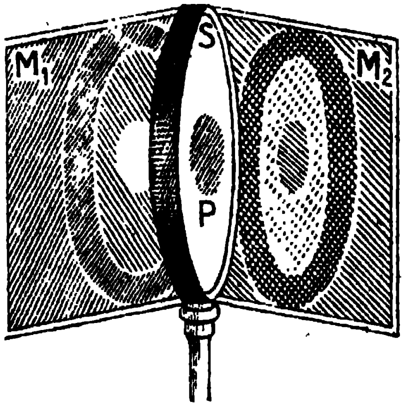


Fig. 154

the spot cannot be distinguished from the rest of the screen. The spot however does not entirely vanish. What we look for is the equality of contrast on both sides between the paper and the spot. To see both sides simultaneously two inclined plane mirrors are placed symmetrically on two sides of the screen (M_1, M_2 ; Fig. 154).

Method. Place two lamps of equal candle power on two sides of the screen P at the same distance. A photometric balance will be obtained. Measure this distance. Let it be d. Keep one lamp fixed at L (Fig. 155) and place three more lamps of exactly the same candle power close together at L_1. This increases the illumination on one side of the screen 4 times. Keeping the four lamps together move them away from L_1 to L_2 until a photometric balance is again obtained. Measure the distance of the group of lamps from the screen of the photometer. Let it be d'. It will be seen that $d'=2d$.

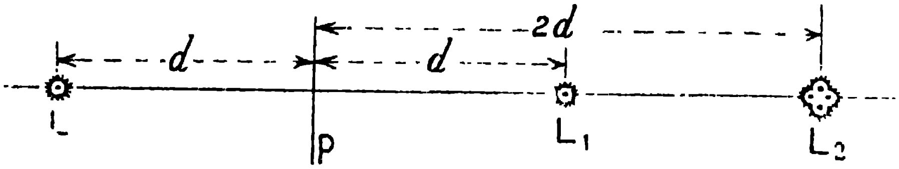


Fig. 155

Repeat the experiment with 9 lamps and 16 lamps. Tabulate to show that when distance is increased the luminous intensity of the source has to be increased as the square of the distance to keep the intensity of illumination constant. Hence intensity of illumination varies inversely as the square of the distance when the luminous intensity of the source is unchanged.

Exercises

1. Define illuminating power and the intensity of illumination. State and explain the units in which they are generally measured.

2. What is a foot-candle? Express it in metre-candles. A 100 c.p. lamp is 5 ft. directly above a point. Calculate the intensity of illumination at the point and express it in foot-candles and also in metre-candles.

(*Ans* : 1 ft-candle=10·76 metre-candles; 4 ft-candles; 43·04 metre-candles).

3. Show by an experiment that the intensity of illumination

due to a point source varies inversely as the square of the distance from the source.

A small 60 c.p. lamp is surrounded by a glass globe which cuts off 25% of the light emitted by the lamp. If a book is held normally 3 ft. below the lamp, what is the intensity of illumination on the book ? (*Ans* : 5 ft-candles).

4. State clearly the principle on which photometry is based. Describe any photometer and explain how you will use it.

5. Two sources, of illuminating powers 144 c.p. and 36 c.p. respectively, are separated by a distance of 90 cm. Determine where between the sources a screen must be placed so that the intensity of illumination on both sides of it may be the same.

(*Ans* : 60 cm from the stronger source).

CHAPTER 22

REFLECTION AT PLANE SURFACES

158. Introductory. When light travelling in a medium falls on a surface separating the first medium from a second, it suffers the following changes:

(i) A part of the light is *thrown back in a definite direction* (or reflected) into the first medium.

(ii) The rest enters the second medium and is said to be refracted.

The proportion of the reflected light to the incident light depends on the nature of the media and on the angle at which the incident light meets the surface. For normal incidence from air to glass about 4% of the incident light is reflected. The amount of reflected light increases as the incidence becomes more oblique. A silvered glass surface reflects 80% or more of the light incident normally on it from air.

In Fig. 156 *MM'* is the surface of separation of two media. A ray *AO* travelling in the first medium is incident at the point *O* on *MM'* and is reflected into the first medium in the direction *OB*. In this case *AO* is called the **incident ray,** *OB* the **reflected ray** and *O* the *point of incidence.* The line *ON* drawn perpendicular to *MM'* at the point *O* is called the *normal at the point of incidence.* $\angle AON$ is the **angle of incidence** and $\angle BON$ the **angle of reflection.** They are equal.

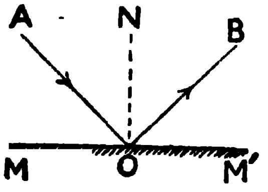
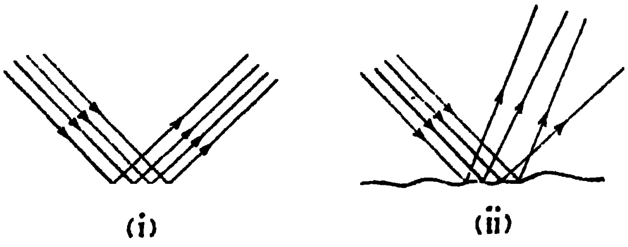


Fig. 156

When the surface *MM'* is plane, a parallel beam of light incident on it will remain parallel even after reflection [Fig. 157 (*i*)]. Reflection in this case is said to the *regular.*



(i) (ii)

Fig. 157

If the surface separating the two media is uneven the normals at different points will point in different directions. At each tiny surface each individual ray will have its angle of reflection equal to its angle of incidence. But the rays that were parallel before reflection will be thrown in random directions [Fig. 158 (*ii*)]. This kind of reflection is called *diffuse reflection* and, loosely, *scattering* by some. A non-luminous object is made visible by the light which it reflects diffusely. Rough surfaces such as walls, papers, the projection screen in a cinema house etc. produce diffuse reflection.

159. Laws of Reflection

Reflection occurs according to two definite laws, which state that:—

(*i*) **The angle of incidence is equal to the angle of reflection.**

(*ii*) **The incident ray, the reflected ray and the normal to the surface at the point of incidence are in one plane.**

Thus in fig. 156, $\angle AON = \angle BON$, and the lines *AO, OB ON* are all in one plane. This plane is called the *plane of incidence*.

160. Experimental Verification of the Laws of Reflection

(i) **By Hartle's optical disc.** The laws of reflection may be verified with the help of Hartle's optical disc. It is a circular disc mounted vertically on a stand (Fig. 158), and can be rotated about an axis through its centre. The disc is graduated in degrees, each quadrant carrying the graduations 0° to 90°. At the centre is fixed a short strip of plane mirror *MM'* with its plane perpendicular to the plane of the disc. The 0°—0° line coincides with the normal to the mirror at the centre of the disc.

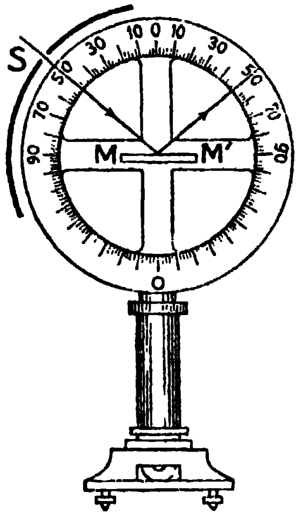


Fig. 158

Light is admitted through a narrow adjustable slit *S* in a screen and falls on the mirror at the centre of the disc. Both the incident and reflected beams graze the surface of the disc, and the angles of incidence and reflection can be directly read off. They are found to be equal.

Reducing the incident beam to a very narrow pencil still grazing the surface of the disc, it is found that the reflected pencil also grazes the surface. They therefore lie in a plane containing the normal.

(ii) By the pin method

Experiment 40. *To verify the laws of reflection by the pin method.*

Apparatus : A thin rectangular strip of mirror, a drawing board, board pins, hair pins, a scale and a protractor.

Procedure: Fix the strip (MM' ; Fig. 159) of mirror vertically on a piece of paper on a horizontal drawing board and mark the reflecting surface on the paper. Fix two pins P_1 and P_2 in front of the mirror. Move about your eye till the images of P_1 and P_2 in the mirror appear in one line. Fix two other pins Q_1 and Q_2 so that they also appear in the same line as the images of P_1 and P_2. Remove the pins and join the points $P_1 P_2$ and $Q_1 Q_2$ See they meet at a point O on MM'. Draw ON perpendicular to MM' at O. Measure $\angle P_1 ON$, the angle of incidence, and also $\angle Q_1 ON$, the angle of reflection. You will find they are equal. Moreover, the incident ray $P_1 P_2$, the reflected ray $Q_1 Q_2$ and the normal ON at the point of incidence are all in the plane of the paper and thus lie in the same plane. Repeat the experiment for at least two other angles of incidence and tabulate the results.

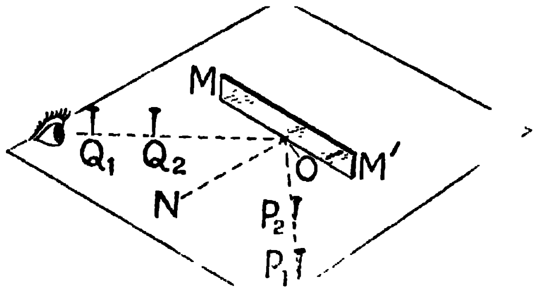


Fig. 159

161. Image formed by a Plane Mirror

To find the position of an image relative to the mirror we may perform the following simple experiment:

Experiment 41. *To determine the position of an image formed by a plane mirror and to verify that the object distance is equal to the image distance.*

Apparatus: A plane mirror, a drawing board, board pins, hair pins, a scale and a protractor.

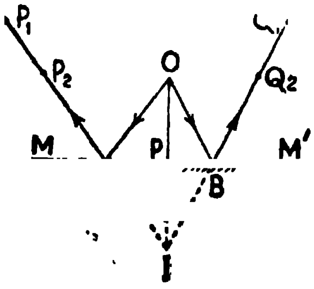


Fig. 160

Method: Mount the mirror vertically on the drawing board and mark the position of its reflecting plane MM' (Fig 160). Fix a pin at O in front of the mirror to serve as object. Fix two pairs of pins, P_1 and P_2, and Q_1 and Q_2, in line with the image of the object at O. Join P_1P_2 and produce it to meet the mirror at A. OA is then the incident ray, and AP_2 P_1 the reflected ray. If Q_1 Q_2 meets the mirror at B, OB is an incident ray and the corresponding reflected ray is BQ_2Q_1. Produce the two reflected rays backwards. Mark their point of intersection I. It is evident that the rays appear to come from I after reflection. Join OI. Measure the angle between this line and MM'. They cut at right angles. Besides, OI is bisected by MM' at P as you will see by measurement. Therefore, *the image is as far behind the mirror as the object is in front of it* .

Repeat the experiment for at least two other positions of the object and tabulate the results.

Theoretical proof. To prove that *image distance=object distance*, we may proceed as follows:

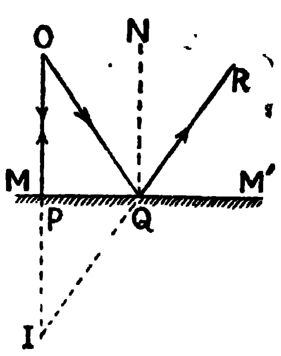


Fig. 161

In fig. 161 let O be a point object in front of the mirror MM'. OP and OQ are two incident rays and their respective reflected rays are PO and QR, OP being incident perpendicularly on the mirror. When the reflected ray PO and QR are produced backwards they intersect at I. The rays thus appear to come from I. I is the image of O.

IO and QN are parallel, each being normal to MM'. Therefore, $\angle POQ = \angle OQN$ and $\angle RQN = \angle PIQ$. But the angle of incidence OQN = the angle of reflection RQN

$$\therefore \quad \angle POQ = \angle PIQ$$

In the $\triangle$s OPQ and IPQ, $\angle POQ = \angle PIQ$ and $\angle OPQ = \angle IPQ$ (both rt. $\angle$s) while the side PQ is common to both.

$\therefore$ The triangles are equal in all respects.

Hence $OP = IP$ or, the object distance = the image distance. We may therefore say that *for reflection at a plane mirror the image lies on the normal drawn from the object to the mirror and is as far behind the mirror as the object is in front of it.*

Virtual image. When rays diverging from a point appear to come from another point the second point is called the *virtual image* of the first. It is clear that the image formed by reflection is a virtual image. A virtual image cannot be received on a screen.

162. Lateral Inversion

Hold the palm of your right hand in front of a plane mirror. The tip of the little finger L (Fig. 162) will have its image at L', L and L' being equidistant from the mirror. Similarly, the tip of the thumb T will have its image at T'. The whole image is similar in size and shape to the right hand, but it is clear from the figure that the image looks like a left hand. Such inversion of right and left on reflection is called lateral inversion.

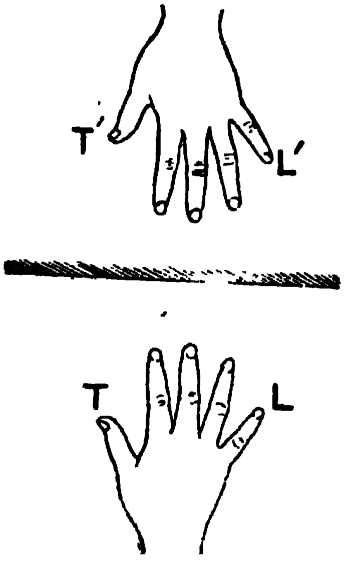


Fig. 162

163. Plane Mirrors and testing Plane-ness

Plane mirrors with which we are mostly familiar are formed by depositing a layer of silver chemically on one face of a plane parallel plate of glass. A polished metal surface reflects the larger part of the light incident on it. Such mirrors are also in use. A clear mercury surface is also a good mirror.

The fact that the object distance is equal to the image distance in the case of a plane mirror enables us to test the plane-ness of a mirror. Look into the mirror for the image of an extended object with straight edges, such as a telegraph post or the bars of a window. If the image shows no distortion the mirror is plane.

164. Simple Periscope

It is a simple device which enables us to see an object behind an obstruction such as a wall. It consists of two plane mirrors, M_1 and M_2 (Fig. 163), fixed at 45° to the main axis of a rather long tube or a frame and has openings against each mirror. In the more improved designs the mirrors can be rotated about horizontal axes. Light from a distant object suffers successive reflections at the mirrors and enters the observer's eye. A simple periscope may be used in a trench to view the happenings outside. It is often seen being used by members of a crowd witnessing a game over the heads of others. Periscopes used in submarines are however much more complicated but the principle is the same.

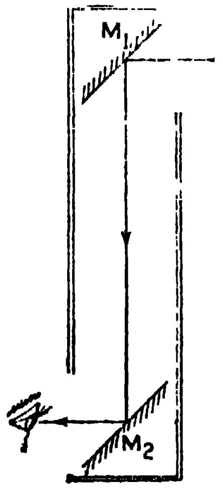


Fig. 163

165. Displacement of the Image in a moving Mirror. When a mirror moves relative to the source the image also moves. Let M_1 be the initial position of the mirror and O that of an object

(Fig. 164*a*). The image will be formed at I_1, where the image distance $I_1 P_1$ is equal to the object distance P_1O. When the mirror moves to M_2, the image is at I_2 where $P_2 I_2 = P_2O$. Therefore the shift of the image $= I_1 I_2 = OI_1 - OI_2 = 2\ (OP_1 - OP_2) = 2P_1P_2$, i.e., *the displacement of the image is double that of the mirror.* When the mirror moves away from the source, the image recedes. The displacement of the mirror and the image are in the same direction.

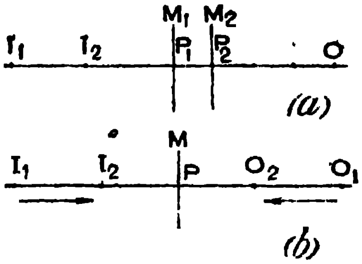


Fig. 164

When the mirror is stationary but the source moves, fig. 164*b* shows that the image moves to the same extent as the source, but its displacement is opposite to that of the source. If v be the velocity of approach of the source towards the mirror, that between the source and the image is $2v$.

166. Effect of Rotation of a Plane Mirror

When a ray of light suffers reflection, it is deviated from its original path. In fig. 165 the incident ray AO would travel along OB' if the surface M_1M_2 were not there. Due to reflection at the surface it travels along OB. The deviation δ which the ray undergoes is

$$\delta = \angle B'OB = 180^\circ - 2i$$

where i is the angle of incidence.

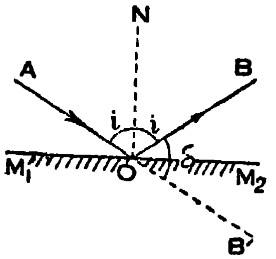


Fig. 165

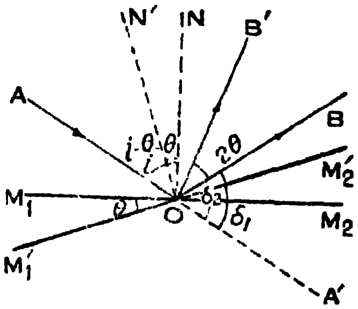


Fig. 166

If a mirror rotates through an angle a Θ ray incident on it from a fixed direction turns through an angle $2\,\Theta$ on reflection (fig. 166). For the deviation of the ray AO when the mirror is in position M_1M_2 is $\angle BOA' = \delta_1 = 180^\circ - 2i$, where i is the angle of incidence AON. When the mirror turns through Θ and is in the position $M_1'M_2'$ the angle of incidence is $AON' = i - \Theta$. Therefore its deviation δ_2 is $180^\circ - 2(i - \Theta)$. The change in the deviation of the reflected ray is

$$\delta_2 - \delta_1 = 2\,\Theta$$

i.e., when the mirror turns through Θ, the image turns throgh 2 Θ,

167. ***The minimum size and position of a mirror in which a person can see his full image.***

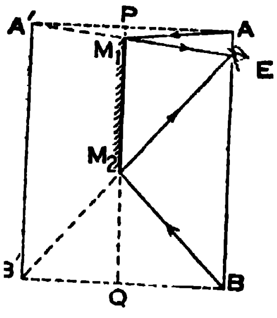


Fig. 167

In fig. 167 AB represents a person and $A'B'$ his image. E is his eye. The plane PQ is equidistant from AB and $A'B'$. The mirror must be in this plane. To find its minimum size and position join $A'E$ and $B'E$. These lines intersect PQ at M_1 and M_2. If the mirror occupies the position M_1M_2 and be of that length, the full image can be seen.

Now in the $\triangle AA'E$, $AP=PA'$ and PM_1 is parallel to AE

Therefore, M_1 is the mid-point of $A'E$.

Similarly in the $\Delta BB'E$, $BQ=QB'$ and QM_2 is parallal to BE.

Therefore, M_2 is the mid.-point of $B'E$.

Now in the $\Delta A'EB'$, since M_1 and M_2 are the mid-points of of $A'E$ and $B'E$ respectively, M_1M_2 is half of $A'B'$

But $AB=A'B.'$

$\therefore\ M_1M_2=\frac{1}{2}AB.$

Hence the minimum length of the mirror is half the height of the person.

168. Inclined Mirrors

(*i*) ***Mirrors inclined at right angle.*** An object P is situated between the mirrors M_1 and M_2 as shown in fig. 168. P_1 is the image of P in M_1; P_2 of P in M_2. P_1' is the image of P_1 in M_2; and P_2' of P_2 in M_1. If the mirrors are accurately at 90°, P_1' and P_2' coincide. Since P_1' and P_2' are formed behind the mirrors, no further image formation is possible. The object and the images being equidistant from O, they lie on a circle with centre O and radius OP.

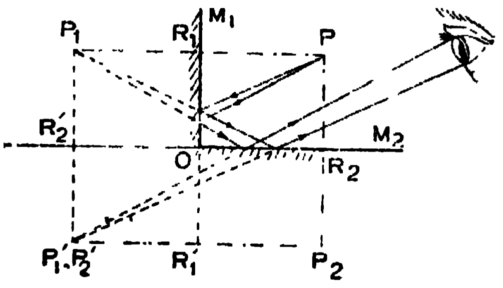


Fig. 168

(*ii*) ***Two mirrors inclined at any angle.*** Fig. 169 represents two mirrors OM_1 and OM_2 inclined at an angle Θ. P is a point object placed between them. P_1' is the image of P in M_1; P_1'', of P_1'in M_2; P_1''', of P_1'' in M_1; etc. A series of images is formed in this way. Similarly, another set of images starting first with reflection in M_2 will be formed. When an image is formed behind both mirrors, no further image formation is possible. It can be easily shown from geometry that the object and all the images are equidistant from O. Hence they lie on the circumference of a circle of radius OP and centre O.

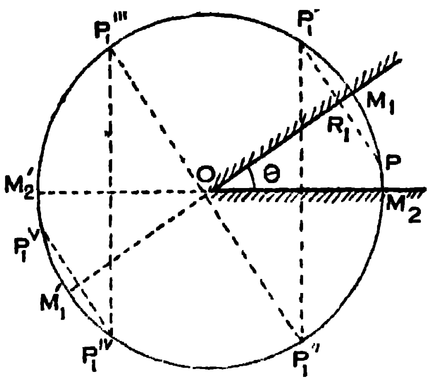


Fig. 169

If Θ is a submultiple of 180°, the total number of images is $\frac{360}{\Theta}-1$. For two mirrors exactly at right angles there will be three images.

Kaleidoscope. It consists generally of three strips of plane mirror, all of the same size, mounted in a cylindrical tube at an angle of 60° to one another. At one end of the tube coloured glass beads are enclosed between two circular glass plates which fit into the tube. When one looks through the observation hole at the other end, beautiful patterns are seen. These are due to the multiple reflections of the coloured beads at the three inclined mirrors.

Exercises

1. State the laws of reflection. How would you verify them?

2. What differences are there in the reflection of light from (*a*) plane mirror, (*b*) a sheet of white blotting paper?

3. Draw a diagram to show how an image is formed in a plane mirror. Prove that the object distance is equal to the image distance.

4. A point object is placed between two mirrors at right angles. Trace the path of rays by which an eye sees an image formed by two successive reflections. Show that the final direction of the beam is parallel to the direction of the incident beam.

5. Two plane mirrors are inclined at 45°. Find by geometrical construction the number of images formed.

6. A man stands at the middle of a room and can fully see a 10′ high wall behind him in a vertical plane mirror placed on the front wall. What is the minimum length of the mirror?

(*Ans*: $\frac{10}{3}$ ft)

7. Describe the construction of a simple periscope. For what purpose is it used?

8. Show that when a plane mirror is rotated through an angle Θ, the reflected beam is turned through an angle 2Θ.

9. (*a*) A man walking towards a stationary mirror at a speed of 2 ft/sec approaches his image at the rate of 4 ft/sec. Explain.

(*b*) If the man is stationary and the mirror is moved towards him at the rate of 2 ft/sec at what speed will his image approach him? (*Ans*: 4 ft/sec.)

CHAPTER 23

REFRACTION AT PLANE SURFACES

169. Some Demonstration Experiments on Refraction

(*i*) Put a coin in an empty bowl and so place your eye that the edge of the bowl just cuts off the coin from view. Hold your head steady and slowly pour water in the bowl. Very soon the coin will come into view. The eye sees here the refracted image of the coin. The bottom of the bowl with the coin over it appears to have been raised. The action may be understood from fig. 170.

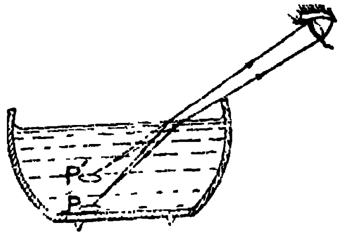


Fig. 170

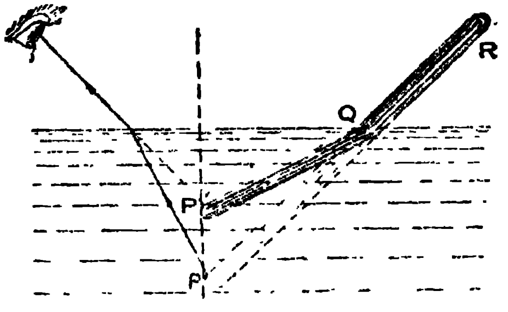


Fig. 171

(*ii*) Dip a rod obliquely in water. It seems to be bent (Fig. 171) at the point where it enters the water. Different portions of the rod under water appear to be raised by different amounts proportionate to their depths. The bending is due to refraction.

(*iii*) Take a glass of water and make it turbid by adding a few drops of milk. Place it in direct sun-light. Take a dark card-board with a small hole in it and hold it in front of the glass so that the hole is a little below the water level (Fig. 172). Note the direction of the beam entering through the hole and passing through water. Now raise the cardboard so that the hole is above the water surface and observe the direction of the beam of sunlight.

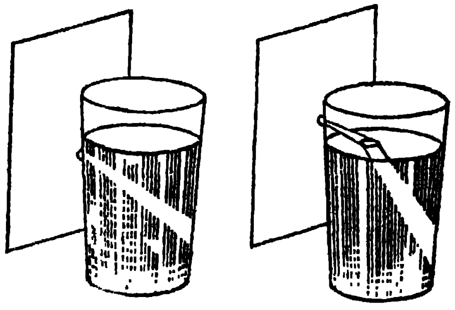


Fig. 172

170. Introductory. When a beam of light travelling in one medium falls on a surface separating it from a second, *a part* of the light enters the second medium. The direction of a ray changes at the point where it enters the second medium. This bending of rays when light enters one medium from another is called ***refraction.***

In fig. 173 let *a* and *b* be two media and *PQ*, their plane of separation. A ray of light *AO* incident on *PQ* at *O* would travel along *AOA′* if the second medium were absent. Due to the presence of the medium *b*, however, the ray will follow a different course, say *OB*. If *NON′* is drawn normal to *PQ* at *O*, ∠*AON* is the *angle of incidence* (*i*) and ∠*BON′*, the *angle of refraction* (*r*). *AO* is the *incident ray* and *OB*, the *refracted ray*. It should be remembered that part of the light incident along *AO* is reflected back into *a* along *OC*.

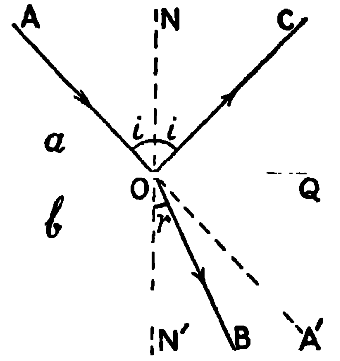


Fig. 173

171. Laws of Refraction. Refraction takes place according to the following laws:

Law I. The incident ray, the refracted ray and the normal at the point of incidence lie in one plane.

This plane is called the *plane of incidence.*

Law II. (*Snell's law*). *For two given media the ratio of the sine of the angle of incidence to the sine of the angle of refraction is a constant for light of a given colour.*

Denoting the constant by the symbol μ we have

$$\frac{\sin i}{\sin r}=\mu$$

$$\text{or} \qquad \sin i=\mu \sin r$$

If $i=0$, it follows that r is also equal to zero. Thus a ray of light incident normally on a surface will, on refraction, travel along the normal in the second medium.

172. Verification of the Laws of Refraction

(i) **By Hartle's optical disc.** Mount a semi-circular slab of glass with its diameter along the 90°—90° line of a Hartle optical disc and the centre of the slab at the centre *O* of the disc (Fig. 174). A narrow parallel pencil of light *AO* strikes the glass slab at *O* and is refracted into the slab. The refracted ray meets the curved glass surface normally and goes out into the air without further bending. Thus though there are two refractions, bending occurs at one surface only. The incident ray and the refracted ray graze the surface of the disc, which also contains the

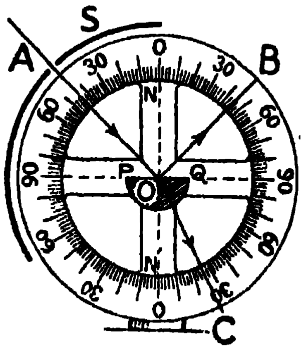


Fig. 174

normal at the point of incidence. They are thus in the same plane. This proves the first law.

To prove the second law the angles of incidence and refraction, i and r, are read off for different positions of the disc and the results tabulated. The ratio $\sin i/\sin r$ is calculated for each position. The values of this ratio for all values of i are found to agree within limits of experimental error.

(ii) By the Pin Method.

Experiment 42. *To verify the laws of refraction.*

Apparatus: A rectangular slab of glass, a drawing board, board pins, hair pins, a protractor, a compass, set squares and a scale.

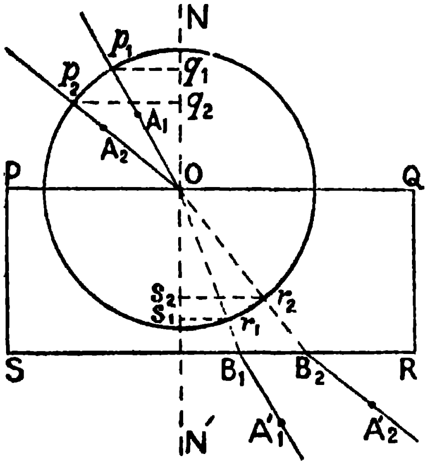


Fig. 175

Procedure: Place the slab of glass on a sheet of paper fixed on the drawing board (Fig. 175). Mark the positions of the sides PQ and RS with pencil lines. Fix a pin O touching the slab, and two other pins A_1 and A_2 at a distance from O. Look through the other face RS of the slab and fix two other pins B_1 and A'_1, so that A_1OB_1 A_1' appear in the same line. Similarly, fix the pins B_2 and A_2' so that they appear in the same line with A_2 O. Remove the glass slab and draw the lines A_1O, OB_1 and B_1A_1' and also the lines A_2O, OB_2 and B_2A_2'. Draw the normal NON' at O.

Draw a circle with O as centre and drop perpendiculars on NON' from the points where the rays A_1O. A_2O, OB_1, OB_2 cut the circle. Measure the lengths p_1q_1, r_1s_1 and p_2q_2, r_2s_2.

OB_2 and OB_1 are the refracted rays corresponding to the incident rays A_2O and A_1O. Therefore, it is to be shown that

$$\frac{\text{sine of the angle of incidence}}{\text{sine of the angle of refraction}}=\frac{\sin A_1ON}{\sin B_1ON'}=\frac{\sin A_2ON}{\sin B_2ON'}$$

$$=\text{a constant, i.e.,}\ \frac{p_1q_1}{r_1s_1}=\frac{p_2q_2}{r_2s_2}=\text{a constant.}$$

Calculate $\frac{p_1q_1}{r_1s_1}$ and $\frac{p_2q_2}{r_2s_2}$. They will be found to be equal.

Alternatively, the angles of incidence and refraction may be measured by a protractor and their sine values obtained from the table and the ratio determined. Note also that the traces of the incident ray, refracted ray and the normal at the point of incidence lie in the same plane.

173. Refractive Index

When a ray of light enters a medium from air, the constant ratio of the sine of the angle of incidence to the sine of the angle of refraction is called the **refractive index** of the medium relative to air. Therefore, refractive index $=\mu=\sin i/\sin r$. The refractive index depends on the nature of the medium but is independent of the value of the angle of incidence. It also depends on the colour of the incident light. The refractive index for red light is less than that for green, blue or violet light. Of two media, the one with the higher refractive index is called the **optically denser** medium. Optical density has nothing to do with physical density. Oil is physically lighter than water, but optically denser.

When light passes from an optically rarer medium into a denser one, the refracted ray bends towards the normal. In the reverse case, i.e., when light passes from a denser to a rarer medium, it bends away from the normal. From fig. 173 you can now say that *a* is the rarer redium and *b* the denser.

Light has the property that *if a ray is reversed in direction it retraces the path along which it travelled.* Referring to fig. 173 again we may, therefore, say that if a ray travelled along *BO* in the medium *b* it would emerge as the ray *OA* in medium *a*. The angles of incidence and refraction are interchanged. *The refractive index of medium 'a' relative to medium 'b' will therefore be the reciprocal of the refractive index of 'b' relative to 'a'.*

Refractive index and velocity of light. Refractive index has an important physical significance. It is the *ratio of the velocity of light in the first medium to that in the second medium.* The velocity of light in air is 186000 miles per second. If the refractive index of glass relative to air is 1·5, velocity of light in glass $=(186000\div 1{\cdot}5)$ miles per second.

Values of some refractive indices relative to air in yellow light

Material	*r.i.*	*Material*	*r.i.*
Glass (Crown)	1·52	Glycerine	1·47
Glass (Flint)	1·65	Turpentine	1·47
Water	1·33	Paraffin	1·44
Ice	1·31	Carbon-disulphide	1·63
Diamond	2·42	Rock salt	1·54

Relative refractive index of two media. Let the refractive indices of two media, *a* and *b* relative to air be μ_a and μ_b respectively. If light enters from '*a*' into '*b*', the refractive index of medium '*b*' relative to medium '*a*' is equal to

$$\frac{\text{Refractive index of medium '}b\text{'}}{\text{Refractive index of medium '}a\text{'}}=\frac{\mu_b}{\mu_a}$$

The r.i. of water is 1·33 and that of glass is 1·52. So for refraction from water to glass, the refractive index $=1{\cdot}52/1{\cdot}33=1{\cdot}14$

174. Geometrical Construction for the Refracted Ray

When a ray *AO* is incident on the surface of separation of air and any other medium, as in fig. 176, we can find the direction of the refracted ray by a simple geometrical construction if μ is known.

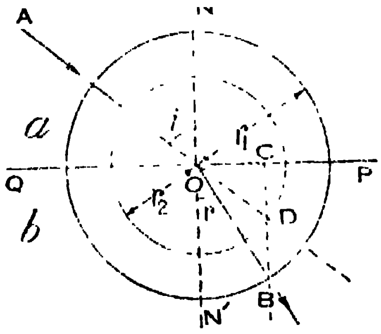


Fig. 176

With *O* as centre draw two circles of radii r_1 and r_2 where $r_2=r_1/\mu$. Let *AO* produced cut the second circle at *D*. Draw *CD* perpendicular on *PQ* and produce it to meet the first circle (in the medium) at *B*. Join *OB*. Then *OB* is the refracted ray.

If *NON′* is perpindicular to *PQ*

$$\sin BON' = \sin OBC = OC/OB = OC/r_1$$

$$\sin AON = \sin ODC = OC/OD = OC/r_2$$

$$\therefore \frac{\sin AON}{\sin BON'}=\frac{OC/r_2}{OC/r_1}=r_1/r_2=\mu$$

Since $\angle AON$ is the angle of incidence, it follows from the above that $\angle BON'$ is the angle of refraction.

175. Virtual Image due to Refraction at a Plane Surface at Normal Incidence.

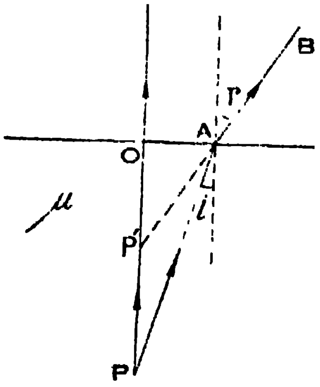


Fig. 177

In fig. 177, let *P* be a point object at a distance $OP=u$ below the plane surface of separation between air and a medium of refractive index μ (say, water or glass). A ray *PA* incident at *A* on the surface at a small angle of incidence *i* will be refracted along *AB* at an angle *r* to the normal at *A*. Produce *BA* backwards to meet *OP* at *P′*. Another ray *PO* incident normally on the surface will be refracted out along the same line. Thus the refracted rays from the medium appear to come from P'. P' is then the ***virtual image*** of P. It is clear from the figure that the the image is raised and is nearer the eye than the object.

Note that μ is the refractive index of the medium relative to air. Therefore, when refraction takes place from the medium into air, $\sin i/\sin r = \dfrac{1}{\mu}$

$$\text{Hence } \frac{1}{\mu}=\frac{\sin i}{\sin r}\quad\frac{\sin APO}{\sin AP'O}=\frac{OA}{PA}\Big/\frac{OA}{P'A}=\frac{P'A}{PA}$$

When the pencil is almost normally incident, A is very near to O and $P'A$ and PA are approximately equal to $P'O$ and PO respectively

$$\therefore\ \frac{1}{\mu}=\frac{P'O}{PO}\ \text{or}\ \mu=\frac{PO}{P'O}$$

Writing $u=PO$=object distance and $v = P'O$=image distance we have $\mu = u/v$ or $v = u/\mu$.

μ for glass or water is greater than unity. Since v gives the apparent position of the object, the apparent thickness of a slab of glass or the depth of a pond will appear to be less than what it is. This accounts for the raising of the coin or the bending of the stick mentioned in § 169.

Experiment 43. *To find the refractive index of glass relative to air.*

Apparatus: Rectangular glass slab, drawing board, pins, paper etc.

Method: Place the glass slab on a sheet of paper on the drawing board and mark its front edge with a pencil line (Fig. 178). Stick a pin at P against its back face, and two pairs of pins, AB and CD, to the front of the block such that ABP appear to be in one line, and so also CDP, when viewed through the glass slab. Remove the slab. Produce AB and CD to meet at P'. Then P' is the apparent position of P when viewed through the glass. The perpendicular distances (OP and OP') of P and P' from the front face of the glass slab are

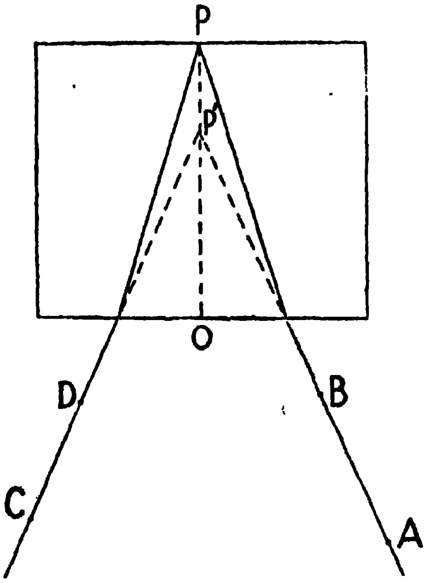


Fig. 178

respectively the real thickness and the apparent thickness of the glass. Measure them. Then

$$\mu = OP/OP' = \ldots\ldots$$

176. Total Internal Reflection

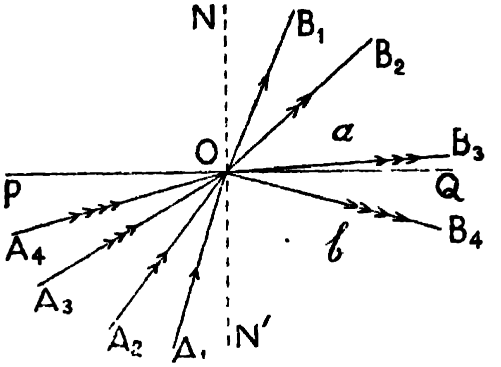


Fig. 179

Consider a ray of light which, travelling in a denser medium '*b*', strikes a surface separating it from a rarer medium '*a*' (air, suppose). Let μ be the refractive index of '*b*' relative to air. As the angle of incidence increases the deviation increases and the refracted ray is bent more and more away from the normal (Fig. 179). The ray A_3O of the figure just grazes the surface on refraction. If the angle of incidence increases still further (as in the case of the ray A_4O) *it is impossible to satisfy Snell's law.* Such a ray is not refracted, but is totally reflected back into the denser medium. This phenomenon is called **total internal reflection** and occurs only for refraction from a denser into a rarer medium. Ray A_3O marks the largest value of the angle of incidence for which refraction is possible, the refracted ray just grazing the surface. The angle of refraction is 90.° The angle of incidence θ_c at this stage satisfies the relation

$$\sin 90°/\sin \theta_c = \mu$$

$$\text{or} \qquad \sin \theta_c = 1/\mu \qquad (176.1)$$

The angle of incidence θ_c in the denser medium for which the refracted ray just grazes the surface of separation is called **critical angle.** Its value depends on the nature of the media and also on the colour of the light. If μ from air to glass is 1·5, the critical angle for refraction from glass to air is given by $\sin \theta_c = 1/1{\cdot}5$, whence $\theta_c = 42°$ (approximately). Between water and air $\sin \theta = 1/1{\cdot}33$ whence $\theta_c = 48\frac{1}{2}°$ nearly.

Between glass and water $\sin \theta_c = 1{\cdot}33/1{\cdot}5$, whence θ_c is about 63°.

177. Some Consequences of Total Reflection

(*i*) Place a test tube obliquely in a beaker of water with the closed end downwards. Allow a horizontal beam of light to fall on the test tube and look from the top in the direction of the emergent rays as shown in fig. 180. The surface of the tube appears bright as a mirror. In looking for an explanation we can neglect the action of the glass wall of the tube, which serves more or less as a parallel plate and does not alter the angle of incidence. If at the water-air surface the light is incident at an angle greater than the critical angle (about 48½°) it will be totally reflected and the surface from which reflection occurs will appear bright. If the tube is made more vertical or is filled with water the brightness disappears, as total reflection ceases.

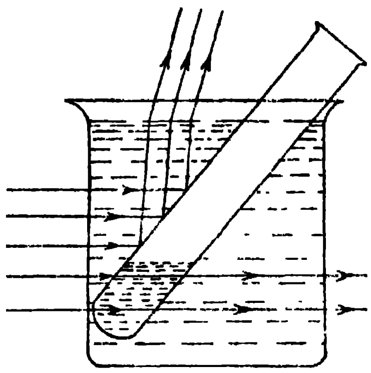


Fig. 180

(*ii*) Make a coarse deposit of soot on a metal ball. Immerse the ball into a beaker of water. Place the beaker in strong, diffuse light. On viewing the ball.its surface will appear shining. The reason lies in the fact that the coarse soot encloses air pockets and light is toally reflected from the water-air surfaces.

(*iii*) When an eye is placed under water rays from an object just above the surface of water enter the eye after refraction at an angle of about 48½° (the critical angle for water-air refraction) with the normal to the surface of water. Objects higher up will enter the eye at smaller angles with the normal than the above (Fig. 181). To an observer inside water all objects above the water surface will appear to lie within a cone of angle about 97°.

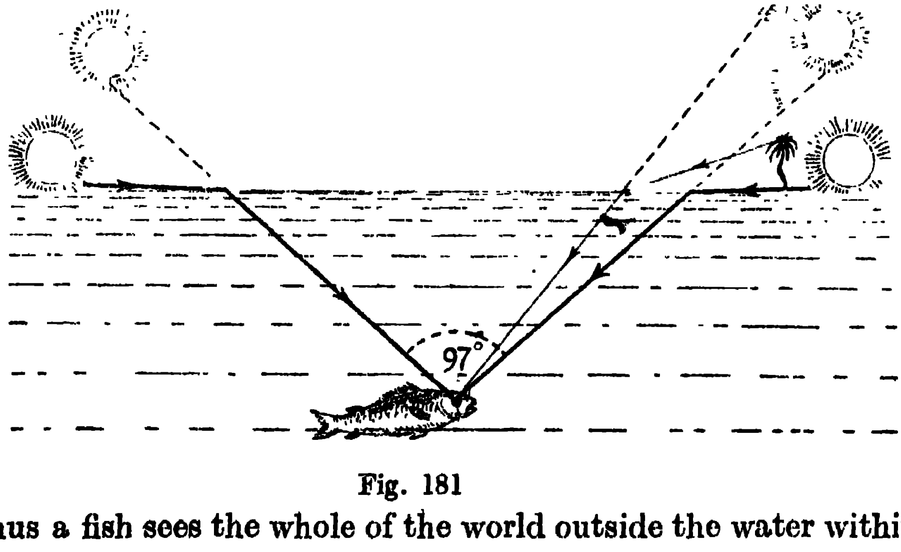


Fig. 181

Thus a fish sees the whole of the world outside the water within a cone of the above angle. The sun describes to it an arc of 97° instead of 180° as to us.

(*iv*) Diamond has a refractive index of 2.4 which makes the critical angle for refraction from diamond to air about 24½°. The faces of a piece of diamond are so cut that the light entering into it suffer total internal reflection repeatedly at different faces and come out through one or two faces only. So these faces appear very glittering. This accounts for the brilliance of diamond and other precious stones.

(*v*) Hold a glass tumbler partially filled with water above the level of the eye and tilt it (Fig. 182). Beyond a certain tilt the surface will appear glittering. Light coming through the bottom has been totally reflected at the top surface. Objects lying at the bottom of the tumbler may be seen reflected at the surface. When properly viewed through one side objects on the other side may also be seen by total reflection at the surface (Fig. 182).

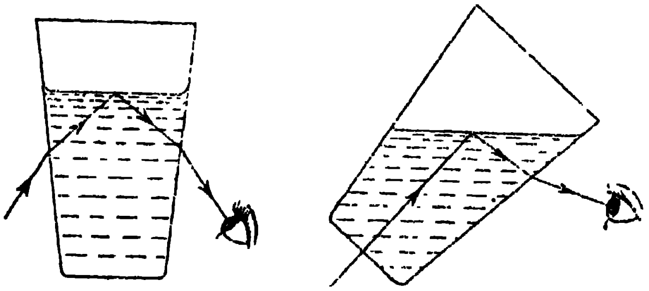


Fig. 182

(*vi*) **Mirage.** It is an optical illusion taking place when the temperature of air above the surface of the earth varies rapidly.

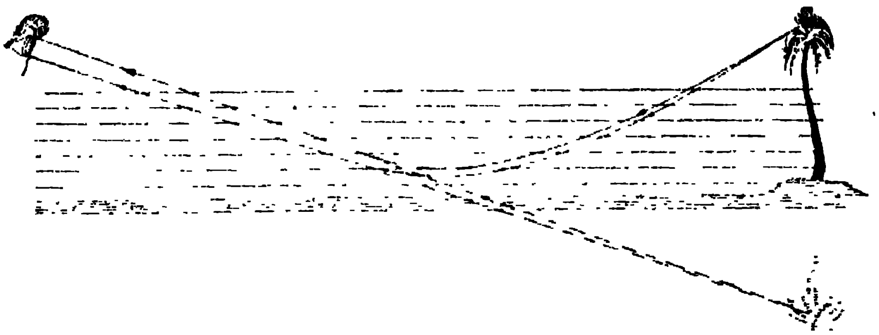


Fig. 183

The refractive index of air (or any gas) diminishes with the increase of temperature. Thus during day time on a desert the sand is very hot and so also the air in contact with it. Upwards the temperature diminishes. Under this condition μ of air is higher above and lower below. Let us assume that layers of air at the same temperature are horizontal (Fig. 183). A ray of light starting from an object above the surface, such as a tree top or a cloud, and moving downwards will enter regions of lower refractive index. It will be bent more and more away from the normal (i.e., the vertical) until at some place it suffers total reflection, and moves upwards. If it is now caught by an eye, the eye will see the object in the direction of the ray produced backwards. To the observer the image may appear as being due to reflection in a pool of water. As the hot air in contact with the sand ascends, the image appears shimmering and unsteady, very much as would happen on a water surface when there is a light breeze. This completes the illusion of reflection in a pool of water. An illusion of this kind is called an *inferior mirage.*

In polar regions the surface of the earth is very cold. The temperature of the air above the surface may increase upwards. A ray of light starting under these conditions from an object on the surface and moving upwards will get into regions of gradually diminishing refractive index. It will be bent more and more away from the normal (Fig 184) until total reflection takes place. The ray is then bent downwards and moves more and more towards the verical. An eye which receives such rays will see the object suspended in air. This illusion is known as a *superior mirage.* The images are more steady and clear than an inferior mirage, as there is no upward motion of the air.

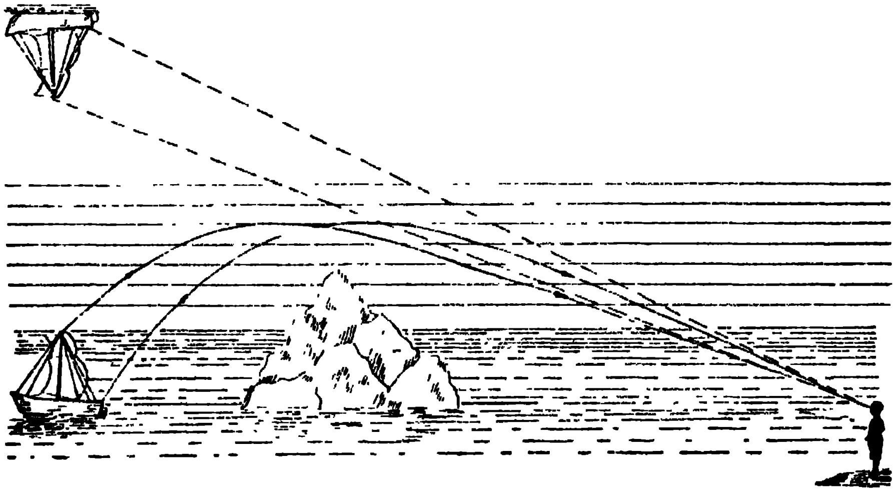


Fig. 184

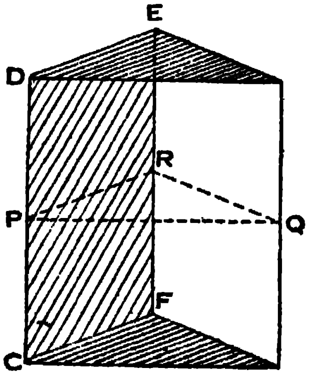


Fig. 185

178. Prisms

For optical purposes a prism may be considered as a portion of a transparent medium two of whose boundaries are plane surfaces inclined to each other at an angle. The line along which the surfaces meet is called the **refracting edge** of the prism (AB of Fig. 185). The surfaces $ABCD$ and $ABFE$ are called the **refracting surfaces**. The angle DAE between them is called the **angle of the**

prism. Any section PQR of the prism by a plane perpendicular to the refracting edge is called a **principal section.**

179. Refraction through a Prism in a Principal Section

Let ABC be a principal section of a prism of angle A (Fig. 186). A ray PQ incident at Q on the refracting face AB is refracted along QR into the prism and out along RS at R. RS is therefore the course of the ray PQ after its passage through the prism. It is clear from the figure that the **emergent ray is bent towards the base of the prism** and is deviated from the original direction. The angle of deviation, δ, is the angle between the incident ray and the emergent ray. δ *increases with the angle A of the prism and is small for thin prisms.* It can be easily shown from geometry that

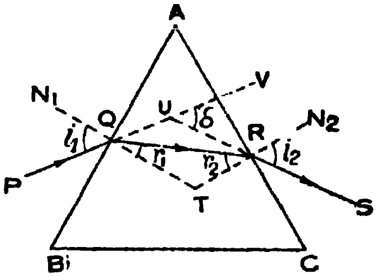


Fig. 186

$$A = r_1 + r_2$$

and $$\delta = i_1 + i_2 - A.$$

These two equations summarize all the geometry of transmission of rays through a prism in a principal section.

180. Minimum Deviation through a Prism

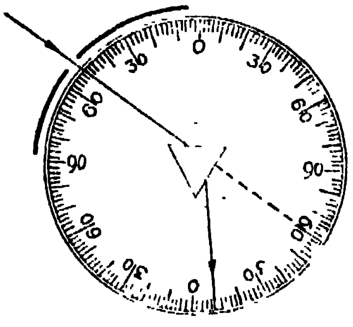


Fig. 187

A prism may be mounted at the centre of a Hartle's disc and illuminated not by white light but by light of some particular colour. This may be obtained by filtering the light from a strong source through red or green glass or by using a sodium flame. For convenience the face on which the light is incident should be parallel to the 90°—90° line and the centre of the prism somewhat below the centre of the disc. By turning the disc i_1 may be varied and corresponding δ read off (Fig. 187).

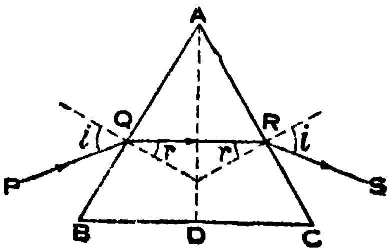


Fig. 187 (a)

If the angle of incidence be gradually decreased, the deviation also gradually decreases until at a particular angle of incidence the deviation attains a minimum value. If the angle of incidence is still further decreased, the deviation, instead of decreasing, increases. There is only one position of the prism for which deviation is a minimum. This position is called the position of minimum deviation of the prism. At the position of minimum deviation $i_1 = i_2$ and $r_1 = r_2$. The ray passes symmetrically through the prism. (Fig. 187 *a*)

Exercises

1. What is refraction? Explain the following phenomena with the help of diagrams :

(*a*) A straight rod dipped partly in water appears to be bent when viewed obliquely.

(*b*) A coin in an empty bowl is invisible to an observer, but when water is gradually poured into the bowl, the coin appears to rise and comes into view.

2. State the laws of refraction. How would you verify them experimentally?

3. Define refractive index. What are the factors on which it depends?

4. Find the missing figures in the following table either geometrically or by calculation.

i (angle of incidence)	r (angle of refraction)	μ (refractive index)
30°	—	1·50
45°	30·3°	—
—	40·6°	1·33.

(*Ans*: 19°·5 ; 1·4 ; 60°)

5. Deduce a relation connecting the real and apparent depths of an object placed in water when viewed normally.

To an observer a coin placed in a basin of water appears to be 4 cm below the top surface. What is the real depth of water in the basin if the refractive index of water is 1.33?

(*Ans* : 5·32 cm)

6. What is total internal reflection? Define critical angle. Find out a relation between the critical angle and the refractive

index of the medium. Calculate the refractive index of glass if the critical angle is 42°.

(*Ans* : 1·5).

7. Describe some consequences of total internal reflection.

8. What is a mirage ? Explain its formation.

9. What is meant by the position of minimum deviation of a prism ? What is meant by saying that in this position light passes through the prism symmetrically ?

10. What do you understand when it is said that a medium is optically denser than another medium ? Arrange water, oil and glass in the order of their optical densities.

11. How is the refractive index of a medium related to the velocity of light in the medium ?

In which of these media does light travel fastest : air, water, oil, glass ?

Does red light travel faster or slower than blue or green light ?

12. You see a fish in water and aim straight at it to spear it. Explain why you cannot succeed.

Why is the external world of a fish confined within a cone of about 97° ?

CHAPTER 24

SPHERICAL MIRRORS AND LENSES

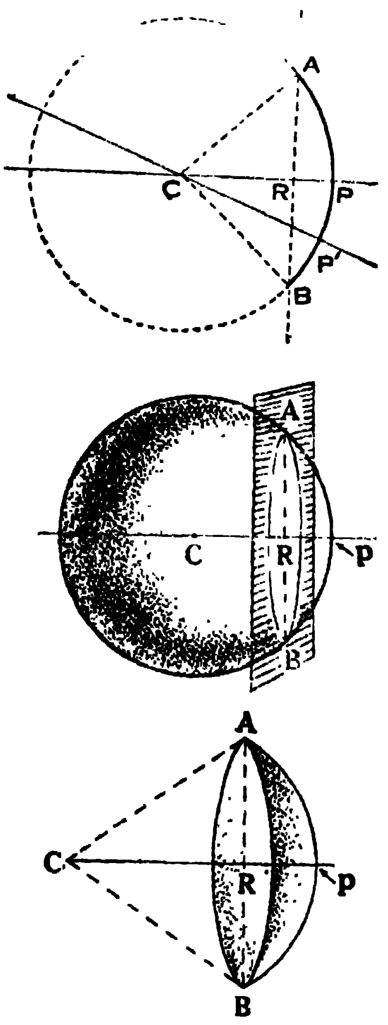


Fig. 188

181. Some Definitions. The surface of a mirror instead of being plane may form part of a surface of regular geometrical shape, such as a sphere, a cylinder or a paraboloid. In this chapter we shall discuss the behaviour of curved mirrors which form part of a spherical surface. A spherical mirror is a part of a spherical shell cut off by a plane (Figs. 188).

When the bulge faces the object the mirror is said to be *convex*. When the hollow faces the object the mirror is *concave*. The centre C (Figs. 188) of the sphere, of which the mirror APB forms a part of the surface, is called the **centre of curvature** of the mirror. The radius CA of the sphere is called the **radius of curvature** of the mirror. The centre of the mirror face (P, Figs. 188) is called the **pole** of the mirror.

The **principal axis** is the line joining the centre of curvature to the centre of the mirror face, that is, the pole of the mirror. Any other line through C (CP', Fig. 188, top) which cuts the mirror is called a *secondary axis*.

182. Image of a Point Object on the Principal Axis

Let O be a luminous point on the axis of a concave mirror APB (Fig. 189), or a convex mirror (Fig. 190) whose centre of curvature is at C. Let a ray OQ be incident on the mirror at a point Q *close to the pole* P. Join CQ. Then CQ is normal to the mirror at Q. OQ is reflected along QR making the same angle i with CQ as OQ does. Let QR (Fig. 189) or QR produced (Fig. 190) cut the axis at I.

Consider the $\triangle OQI$. The angle at Q is bisected by QC internally in fig. 189 and externally in fig. 190. From elementary geometry we know that the bisector of the vertical angle divides the base in the ratio of the sides. Therefore $CO/CI = QO/QI$.

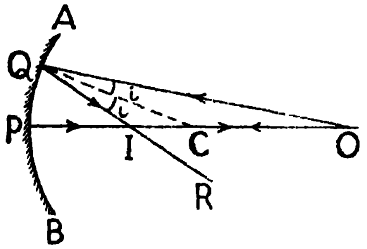


Fig. 189

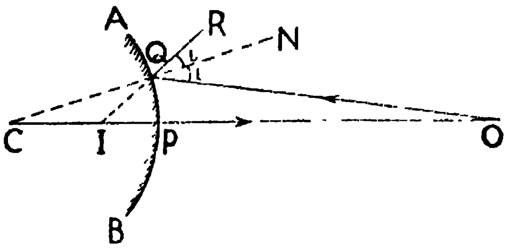


Fig. 190

Now since Q is very close to the pole P, QO is very nearly equal to PO and QI to PI.

$$\therefore \quad CO/CI = PO/PI$$

Therefore I is a fixed point. Thus all rays starting from O and incident at points close to P will pass through I when the mirror in concave. They will appear to come from I when the mirror is convex. I is the image of O.

183. Real and Virtual Images

When rays diverging from a point either meet at or appear to come from another point, the second point is called the **image** of the first.

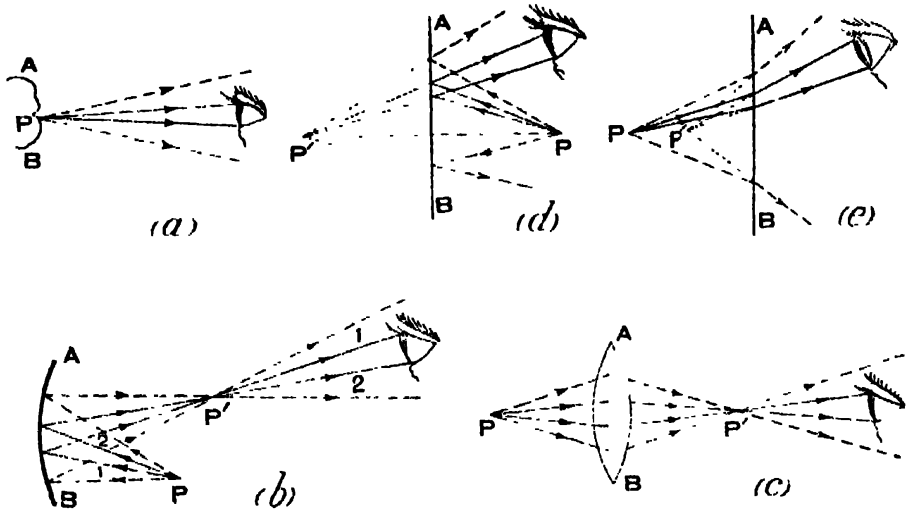


Fig. 191

When rays starting from an object point meet at another point after reflection or refraction, the latter point is called the **real image** of the former.

If rays starting from an object point appear, after reflection or refraction, to come from another point, the latter point is called the **virtual image** of the former.

Fig. 191 (*a*) shows how an object point P is directly seen by the eye. (*b*) and (*c*) show how a real image of P is formed at P' by reflection at a concave mirror and refraction at a lens respectively. (*d*) and (*e*) illustrate formation of virtual images by reflection and refraction.

In fig. 189, the image is real. In fig. 190 it is virtual.

184. Convention of Signs

In problems of reflection at a spherical surface the distance concerned (i.e., the distance of the object, the image, the centre of curvature etc.) are measured from the pole of the mirror. Sometimes these points are to the right, at other times to the left of the mirror. To indicate clearly the position of the points with respect to the mirror their distances from the mirror are taken as positive or negative according to their position. Various conventions regarding the sign of these distances are in use. In this book we have adopted the following conventions :

All distances are measured from the pole of the mirror. Distances measured in the direction opposite to that in which light proceeds from the source towards the mirror are taken as positive. The direction in which light proceeds from the source towards the mirror is taken as negative.

In figures 189 and 190, where the object is at O the object distance PO to be measured from the pole P to O is positive, as this direction is opposite to that in which light moves in proceeding from the object to the mirror. In fig. 190 the image distance PI, which we always measure from the pole, is negative.

185. Position of the Image of a Point on the Principal Axis

In § 182 we have seen that I, the image of O, is a fixed point whose position is defined by the relation

$$CO/CI = PO/PI$$

Let the object distance $PO=u$, the image distance $PI=v$ and the radius of curvature $PC=r$. For the concave mirror (Fig. 189) all these quantities are positive according to the convention of signs we have adopted.

$\therefore$ For the concave mirror,

$$\frac{u}{v}=\frac{PO}{PI}=\frac{CO}{CI}=\frac{PO-PC}{PC-PI}=\frac{u-r}{r-v}$$

or $$u(r-v)=v(u-r)$$

or $$ur+vr=2uv$$

Dividing by uvr we have

$$\frac{1}{v}+\frac{1}{u}=\frac{2}{r} \qquad (185.1)$$

The same formula applies to convex mirrors, but r will be negative according to the convention of signs.

Conjugate points. Pairs of points on the principal axis of a spherical mirror whose distances u and v from the mirror satisfy the equation 185.1 are called **conjugate points**. It will be seen from the equation that the values of u and v may be interchanged without violating the equation. Referring to the figs. 189 and 190 this means that if we place an object at I, its image will be formed at O. O and I are conjugate points. In other words *a pair of conjugate points are such that if an object is placed at any one of them the image will be formed at the other.*

186. Principal Focus and Focal Length of Spherical Mirror

A pencil of rays parallel to the principal axis of a spherical mirror comes to a focus at a point midway between the mirror and its centre of curvature. This point is called the **principal focus** of the mirror and its distance from the pole is called the **focal length** of the mirror. In this case $u=\infty$, which gives $v=r/2$. The focal length f is half the radius of curvature

$$f=r/2 \qquad (186.1)$$

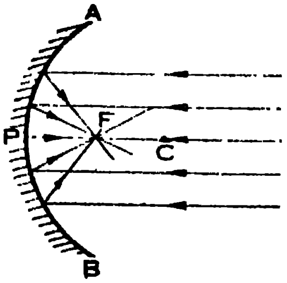


Fig. 192

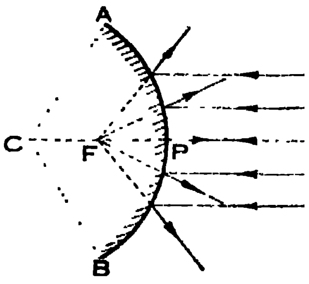


Fig. 193

In the case of a concave mirror all rays of a pencil parallel to the principal axis pass after reflection (Fig. 192) through the focus F. The focus in this case is called a *real* focus.

In the case of a convex mirror r is negative and the focus F (Fig. 193) lies behind the mirror. A bundle of parallel rays, therefore, appears after reflection to come from F. The focus is *virtual.*

A plane perpendicular to the principal axis and passing through the principal focus is called the **focal plane.**

In view of the relation $f=r/2$, Eq. 185.1 for spherical mirrors may be written as

$$\frac{1}{u}+\frac{1}{v}=\frac{1}{f} \qquad (186.2)$$

which is the more usual form.

Point object at the principal focus. If a point object is placed at the principal focus of a concave mirror, rays starting from it will, after reflection at the mirror, be rendered parallel to the principal axis (Fig. 194). In the case of a convex mirror, rays proceeding towards the principal focus will be rendered parallel to the principal axis after reflection (Fig. 195).

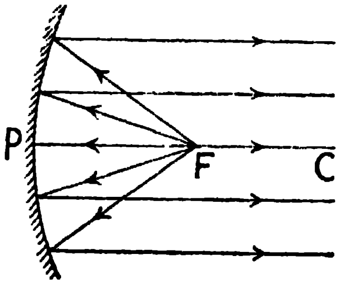


Fig. 194

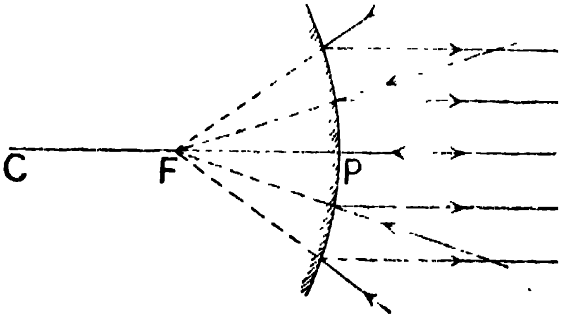


Fig. 195

187. Construction for the Image of a finite Object. Consider an object of finite size placed on the principal axis of a mirror. A simple geometrical construction will give the size and position of the image. In making such a construction we have to remember that

(1) *A ray of light parallel to the principal axis of a concave mirror passes through the principal focus after reflection. In the case of a convex mirror it appears to come from that point.*

(2) *A ray passing through the centre of curvature of a concave mirror (or moving towards the same point in the case of a convex mirror) falls normally on the mirror and retraces its path.*

(3) *A ray passing through the principal focus of a concave mirror (or proceeding towards this point in the case of a convex mirror) will, after reflection, proceed parallel to the principal axis.*

To find the position of the image of a point by geometrical construction, we may draw from the object *any two* of the three rays specified above. Their point of intersection (real or virtual) after reflection gives the position of the image. All rays starting from the point object meet at the same point.

In what follows we have generally used the rays (1) and (2) mentioned above, though any two may be taken.

Construction for the image of an extended object. Let *PQ* be a straight object on the axis of a spherical mirror and perpendicular to it. *C* and *F* are respectively the centre of curvature and the principal focus of the mirror. To find the position and size of the image draw from Q a ray QA parallel to the axis. For a concave mirror (Fig. 196) it passes through F, while for a convex mirror (Fig. 197) it appears to come from F after reflection. Another ray from Q passing through C (Fig. 196) or moving towards C (Fig. 197) is reflected back along the same path. The point of intersection Q' of these two rays is the image of Q. From Q' draw $Q'P'$ perpendicular on the principal axis. $P'Q'$ is then the image of PQ. If the object extends on the other side of the principal axis a similar construction may be made for the image of the other end of the object.

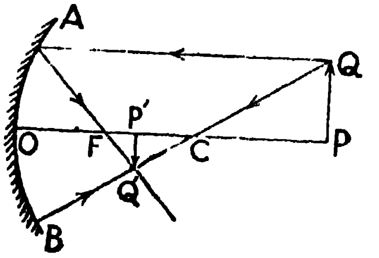


Fig. 196

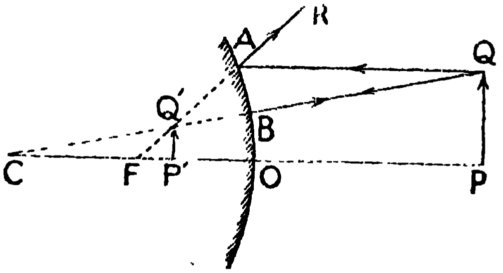


Fig. 197

In fig. 196 the image is real and inverted, while in fig. 197 it is virtual and erect.

188. Magnification. The ratio of the size I of the image to the size O of the object is called **magnification.** Denoting it by m and referring to fig. 196 or 197 we have

$$m = \frac{I}{O} = \frac{P'Q'}{PQ}$$

Writing object distance $OP=u$, image distance $OP'=v$, radius of curvature $OC=r$, we have from the similar triangles CPQ and $CP'Q'$

$$\frac{P'Q'}{PQ} = \frac{CP'}{CP} = \frac{r-v}{u-r}$$

From the formula $\frac{1}{u} + \frac{1}{v} = \frac{2}{r}$

we have $\frac{1}{v} - \frac{1}{r} = \frac{1}{r} - \frac{1}{u}$

or $\frac{r-v}{vr} = \frac{u-r}{ur}$

or $\frac{r-v}{u-r} = \frac{v}{u}$

$$\therefore \quad m = \frac{P'Q'}{PQ} = \frac{v}{u}$$

or, magnification $= \frac{\text{image distance}}{\text{object distance}}$

When $m > 1$, the image is said to be *magnified*; when $m < 1$, the image is said to be *reduced.*

189. Some uses of Spherical Mirrors

(*i*) **The phantom doll.** An interesting optical illusion may be produced with the help of a concave mirror of relatively large radius of curvature. We have seen that a concave mirror forms a real magnified and inverted image of an object placed between its centre of curvature and focus.

If a small doll, or any other convenient object, be placed in an inverted position below the principal axis of a concave mirror between its focus and the centre of curvature, a real magnified image of the object will be formed by the mirror (Fig. 198). An eye at a distance of more than 10″ (the least distance of distinct vision) from the image and suitably placed to receive the rays forming the image will see a doll suspended in air. Viewed from any other angle where the reflected rays cannot reach the eye the doll cannot be seen. When arranged in a dark room with the actual doll screened from direct view and illuminated by hidden light the appearance is very realistic.

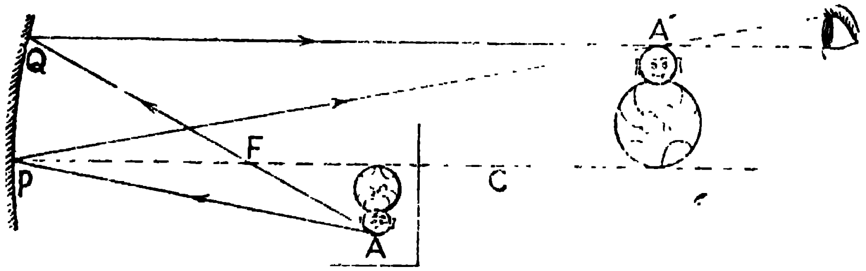


Fig. 198

(*ii*) **Shaving mirror.** A concave mirror forms an erect, magnified and virtual image of an object placed between its focus and pole (Fig. 199). If a man brings his face close enough to a concave mirror he sees a magnified virtual image of his face in the mirror. For convenience of vision the distance of the image should be greater than 10″ from his eye, because a normal eye finds it difficult to see things clearly when nearer than this.

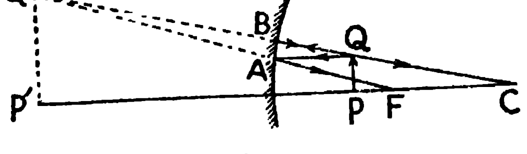


Fig. 199

(iii) **View finder of an automobile.** To get a view of the traffic in the rear the driver of an automobile often uses a convex mirror projecting outside the car. All real objects in front of the mirror (i.e., to the rear of the car) form erect, diminished and virtual images lying between the focus and the pole of the mirror. The driver can therefore get a view of everything that lies to his rear by simply looking into the mirror.

Fig. 200 shows an object PQ placed in three different positions marked I, II and III on the principal axis of a convex mirror.

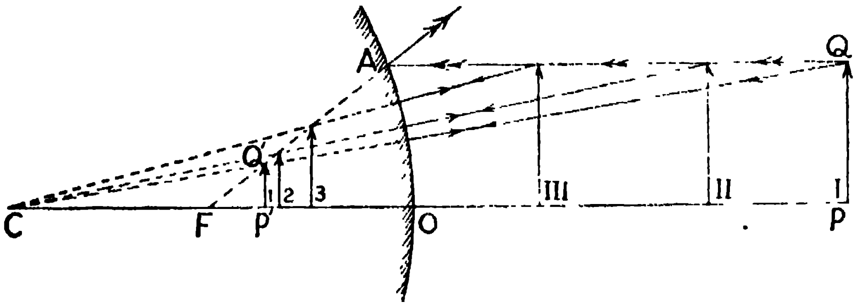


Fig. 200

The corresponding images are found to be formed at positions marked 1, 2 and 3, all between the pole and the focus. All the images are virtual, erect and diminished in size.

Experiment 44. *To determine the focal length of a concave mirror by u-v method.*

Apparatus : An optical bench, a candle flame, paper screen and concave mirror.

Method : Place a candle flame close to the principal axis of a concave mirror, preferably between its focus and the centre of curvature. Keeping the positions of the mirror and the candle fixed, adjust the position of the paper screen so that a sharp image is formed on it . Find the distance between the flame and the mirror. This is the object distance u. Similarly find the image distance v. The relation $\frac{1}{u}+\frac{1}{v}=\frac{1}{f}$ gives the focal length.

Repeat the experiment for various positions of the object and find f from each observation. Tabulate as follows :

No. of observations	*Object distance u in cm*	*Image distance v in cm*	*focal length $f=\frac{uv}{u+v}$ cm*	*Average f in cm*
1				
2				
3				

190. Lenses

A *lens* is a portion of a refracting medium bounded by two surfaces of definite geometrical form. The surfaces may be parts of spheres or cylinders, the former being the more common. One of the surfaces may be plane. When both the bounding surfaces are spherical the line joining the centres of curvacure of the two surfaces is called the **principal axis.**

Lenses are generally divided into two classes, viz., (i) **convex** and (ii) **concave.** Convex or converging lenses are thicker at the middle than at the edges. They may be (*a*) *double-convex,* in which the centres of curvature lie on two sides of the lens, (*b*) *convex-meniscus,* in which both centres lie on the same side of the lens or (*c*) *plano-convex,* where one surface is plane. The forms are shown in fig. 201

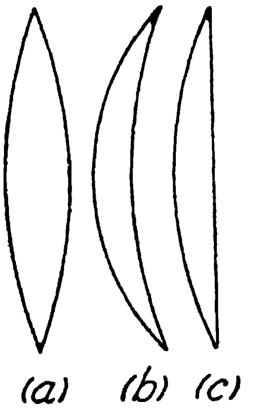


Fig. 201

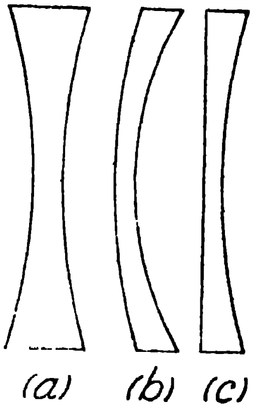


Fig. 202

Concave or diverging lenses are thinner at the middle than at the edge and may be (*a*) *double-concave,* (*b*) *concave-meniscus* or (*c*) *plano-concave* as shown in fig. 202.

191. A Lens is like a Set of Prisms

The action of a lens on a pencil of rays passing through it depends on the nature of the lens as well as the refractive index of its material with respect to the surrounding medium.

The convergent or the divergent action of a lens may be easily understood when we compare a lens with a prism. A lens may be looked upon as a combination of a series of prisms (Fig. 203 and 204) whose angles increase continuously from the centre towards the margin of the lens. In the case of a convex lens, the bases of the prisms are towards the centre of the lens (Fig. 203). A ray passing through such a prism will be deviated towards the base and the deviation will increase as the angle of the prism increases (vide § 179). Thus the rays of a parallel bundle incident on a convex lens are brought nearer together and may be made to converge at a single point after refraction. In the case of a concave lens the prisms have their bases turned away from the centre. Due to refraction towards the base the rays of a parallel bundle are made divergent on passing through a concave lens (Fig. 204).

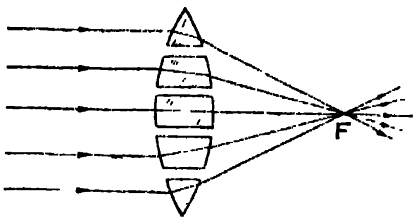


Fig. 203

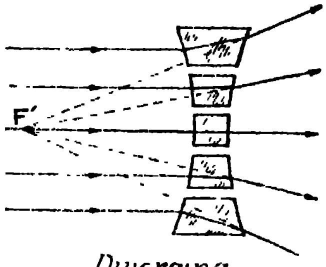


Fig. 204

192. Optical Centre of a Lens

There is a point on the principal axis of a lens such that a ray passing through the point suffers no deviation, but emerges from the lens in a direction parallel to its direction of incidence on the lens. This point is called the *optical centre of the lens.*

In the elementary theory of lenses, the thickness of a lens is not taken into consideration. Such lenses are called ***thin lenses.*** In the case of a thin lens, therefore, any point on the principal axis lying within the lens may be taken as the optical centre for purposes of geometrical construction.

In our diagrams we shall show the outline of the lens to indicate its character, convex or concave. But refraction will be shown to take place at a plane indicative of the position of the thin lens.

Sign convention. In the case of a lens, distances are measured from the optical centre. Distances measured against the direction in which light travels are taken as positive; those in the same direction are negative. The convention is the same as for spherical mirrors.

193. Principal Foci and Focal Length

There is a point on the principal axis of a lens such that a bundle of rays starting from the point will, after refraction by a convex lens, be rendered parallel to the axis (Fig. 205).

For a concave lens, a bundle of rays proceeding towards this point (Fig. 206) will be rendered parallel after refraction. This point is called the **first principal focus.** The distance of this point from the optical centre is called the **first focal length** (f_1). For a convex lens f_1 is positive; for a concave lens, negative.

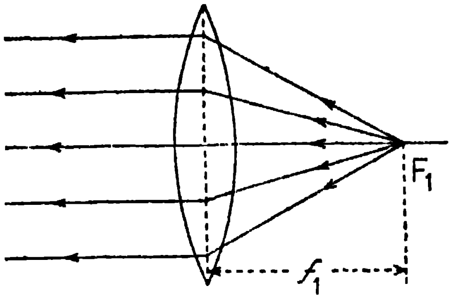


Fig. 205

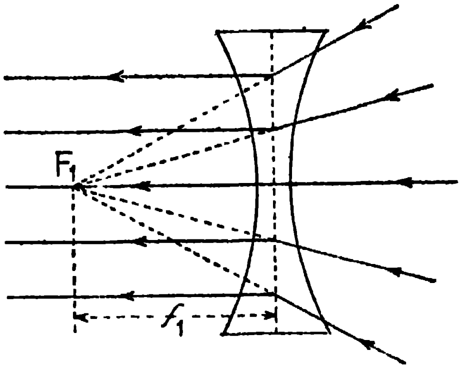


Fig. 206

A bundle of rays parallel to the principal axis will after refraction through a convex lens converge to a point on the principal axis (Fig. 207). For a concave lens it will appear to diverge from a point on the principal axis (Fig. 208). This point is called the **second principal focus** of the lens. The distance of this point from the optical centre is called the **second focal length** (f_2). For a convex lens f_2 is negative; for a concave lens it is positive.

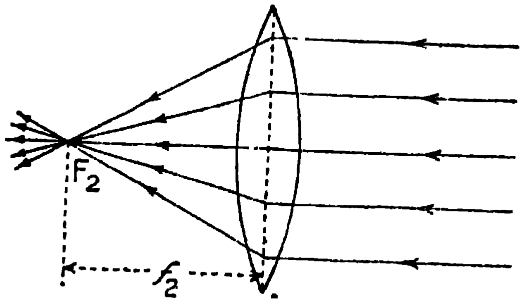


Fig. 207

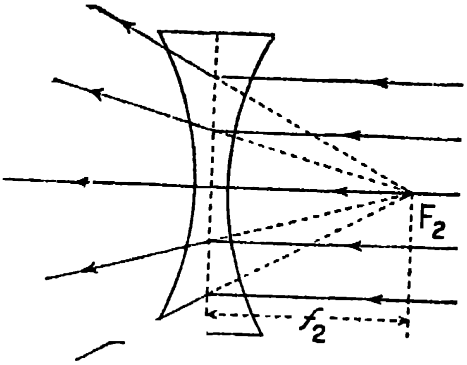


Fig. 208

The focal points are fixed with respect to a lens. For a lens placed in air the points are equidistant from the centre of the lens. *By the term* **focal length** *of a lens is understood the distance of the second principal focus from the centre of the lens. Conventionally for convex lenses the focal length is taken as negative* and *for concave lenses, positive.*

A plane through a focus perpendicular to the principal axis is called a *focal plane.*

194. Graphical Construction for Image

In finding the position of image of an extended object lying perpendicular to the principal axis of a lens we have to remember any two of the following properties of a lens:

(1) *A ray passing through the first principal focus of a convex lens, or proceeding towards the same point of a concave lens, will after refraction travel parallel to the principal axis.*

(2) *A ray of light starting from a point on the object and travelling parallel to the principal axis will after refraction in a convex lens pass through, or, in a concave lens, appear to come from, the second principal focus.*

(3) *A ray passing through the optical centre of a lens emerges without deviation.*

These rays after refraction either meet at a point or appear to come from one and the same point (i.e., they meet when produced backwards). In the former case, the point at which they meet is called the *real image* of the object point. In the latter case, the point of backward intersection of the rays is called the *virtual image* of the object point.

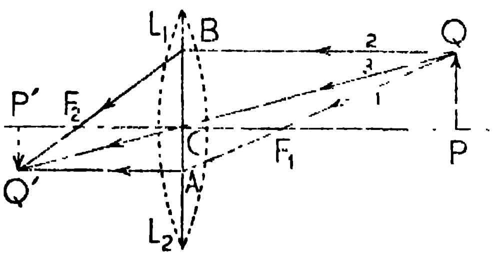


Fig. 209

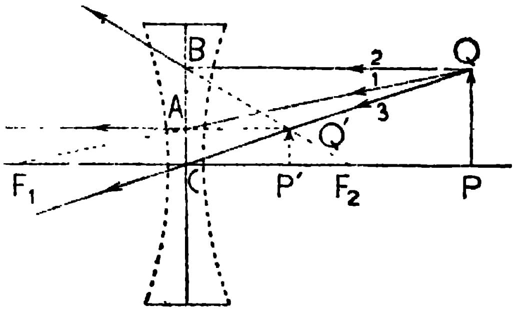


Fig. 210

The above facts are illustrated in figs. 209 and 210. The object PQ stands perpendicular to the axis of a lens of which F_1 and F_2 are the first and the second principal foci respectively and C the optical centre. The ray QA passes through F_1, QB through F_2 and QC through the optical centre. They are respectively marked 1, 2 and 3 in agreement with the numbering of the rays mentioned in italics at the beginning of this article.

195. Derivation of the Lens Formula

From the similar $\triangle^s$ CPQ, $CP'Q'$ of figs. 209 and 210

we have $$PQ/P'Q'=CP/CP' \qquad \text{(A)}$$

Again, from the $\triangle^s$ CBF_2 and $P'Q'F_2$

$$CB/P'Q'=CF_2/F_2P' \qquad \text{(B)}$$

But since $CB=PQ$, the right hand sides of (A) and (B) are equal

$$\therefore \quad CP/CP'=CF_2/F_2P' \qquad (C)$$

Let us put the object distance $CP=u$, the image distance $CP'=v$ and $CF_2=f$, the second focal length.

For the convex lens (Fig. 209) v and f are negative

Remembering the sign convention, we get

$$F_2P'=CP'-CF_2=(-v)-(-f)=f-v.$$

Hence from (C) $$\frac{u}{-v}=\frac{-f}{f-v}$$

or $$uf-uv=vf.$$

Dividing both sides by uvf

we have $$\frac{1}{v}-\frac{1}{u}=\frac{1}{f}$$

The same formula will apply to concave lenses, but f should be taken as positive.

Convex lens forming real image. In the case of a convex lens forming a real image, both f and v are negative. In such a case we may write the working formula as

$$\frac{1}{u}+\frac{1}{v}=\frac{1}{f}$$

where u, v and f represent the *magnitudes* of the corresponding quantities. They are no longer to be taken with any special sign.

So long as u is greater than f, a convex lens forms a real image.

196. Magnification. *Magnification is defined as the ratio of the size of the image to the size of the object.* From the similar triangles CPQ and $CP'Q'$ (Figs. 209 and 210) we have

$$\textit{magnification}\ \mathbf{m}=\frac{P'Q'}{PQ}=\frac{CP'}{CP}=\frac{v}{u}$$

In fig. 209 for the convex lens, u is positive but v is negative and the image is inverted. In fig. 210 for the concave lens, both u and v are positive and the image is erect. *A negative sign in the expression for magnification thus represents an inverted image while a positive sign means an erect image.*

197. Power of a lens. The reciprocal of the focal length of a lens is called its **power**. It is measured in **dioptres**. *One dioptre is the power of a lens of focal length one metre.* To get the power in dioptres, express the focal length in metres and take the reciprocal, i.e.,

$$\textit{Power in dioptres}=\frac{1}{\textit{focal length in metres}}=\frac{100}{\textit{focal length in cm}}$$

Unlike the convention used in text books of physics, opticians take the power of a convergent lens to be positive and of a divergent lens, negative. The power of a lens is a measure of *its ability* *to produce convergence or divergence.* A lens of larger power, i.e., of shorter focal length, produces stronger convergence if it is convex and stronger divergence if it is concave.

When two thin lenses are in contact, the power of the combination is the algebraic sum of the powers of the components.

198. Images formed by a Convex Lens for different Positions of an Object

(*i*) For an object at infinity (a great distance), the image is formed in the focal plane. The image is extremely reduced in size and is inverted (Fig. 211),

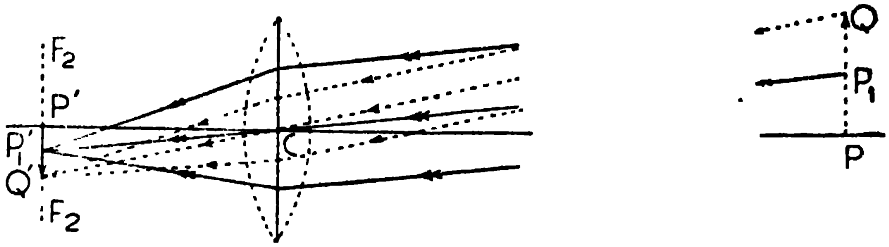


Fig. 211

This provides a method of measuring the focal length of a lens. On a white screen, form with a lens a sharp image of an object about 60 ft or more in distance. The distance between the lens and the screen is the focal length.

(*ii*) For an object at a distance of twice the focal length, an inverted image of the same size as the object is formed at the same distance on the other side of the lens (Fig. 212)

(*iii*) For an object between the focus and a distance of twice the focal length, the image is real, inverted and magnified (Fig. 213)

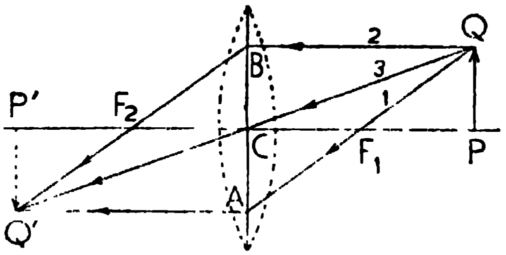


Fig. 212

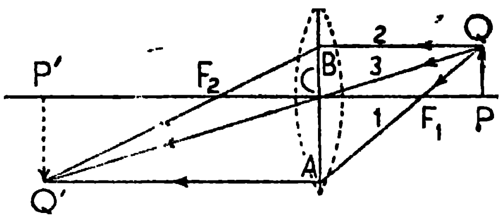


Fig. 213

(*iv*) For an object in the first principal focal plane, the image will appear to an observer to lie at infinity (a great distance). It is virtual and erect (Fig. 214). In one method of using a magnifying glass the object is placed in the first principal focal plane.

(*v*) When the object lies between the focus and the pole, the image is virtual, erect and magnified (Fig. 215). This gives another method of using a magnifying glass (Sec. § 204).

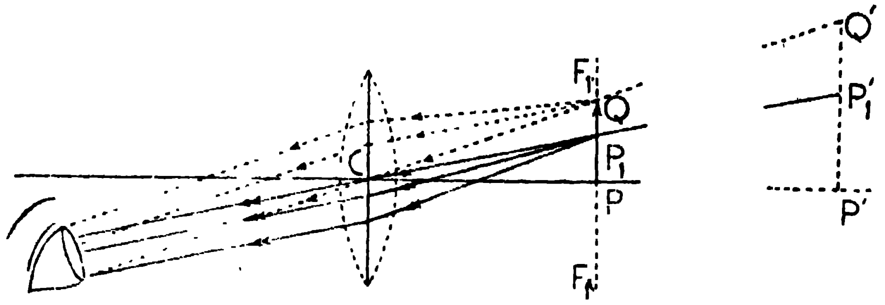


Fig. 214

Images formed by a concave lens

For all positions of an object, the image lies between the focus

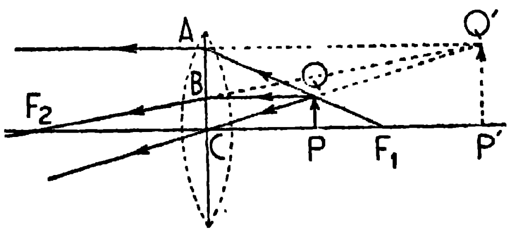


Fig. 215

and the pole. The image is virtual, erect and reduced. Fig. 216 shows three positions marked I, II and III of the object PQ and the corresponding positions 1, 2 and 3 of the image $P'Q'$.

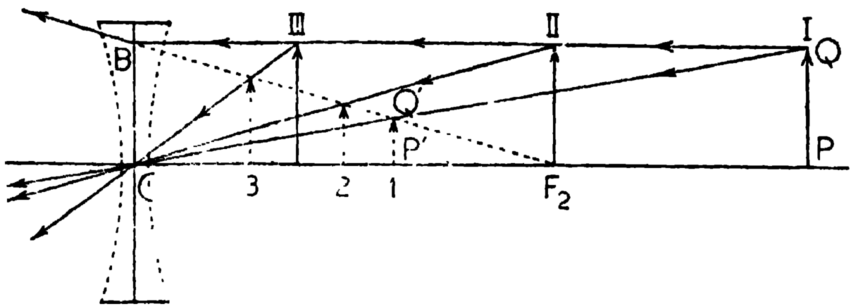


Fig. 216

199. Determination of Focal Length of a Convex Lens

Experiment 45. *To determine the focal length of a convex lens by U-V method.*

Apparatus: An optical bench, an electric lamp in a lamp house, a convex lens and a paper screen.

Method: Mount a cross wire on a vertical upright provided in the optical bench. Place it at one end of the bench. Place an electric lamp behind it so as to illuminate it properly. The paper screen is to be placed on the other end of the bench. Interpose the convex lens between the crosswire and the screen. Adjust the height of the uprights so that the intersection of the cross wire, and centre of the paper screen lie along the axis of the lens. Move the lens slowly until a sharp image of the crosswire is received on the screen. Measure the object distance u and the image distance v. The focal length f of the lens can be calculated from the formula $\frac{1}{u}+\frac{1}{v}=\frac{1}{f}$ (*Vide* § 195).

Repeat the experiment moving the lens away from the crosswire by a centimetre or two. Find f from each observation. Tabulate as follows:

No. of observations	*Object distance* (u) *in cm*	*Image distance* (v) *in cm*	*Focal length* $f=\frac{vu}{u+v}$ *cm*	*Average* f *in cm*
1				
2				
3				

Experiment 46. *To determine the focal length of a convex lens by plane mirror method.*

Apparatus. A plane mirror, a pin with a sharp point, an adjustable clamp, a convex lens and a scale.

Principle of the method. When a point object is placed at the first principal focus of a convex lens, a beam of rays starting from it will be rendered parallel to the principal axis after refraction through the lens. If a plane mirror be placed behind the lens L (Fig. 217) with its plane perpendicular to the principal axis of the lens, the rays will retrace their paths after reflection at the mirror. The lens will then bring these rays to a focus at the same position as the object.

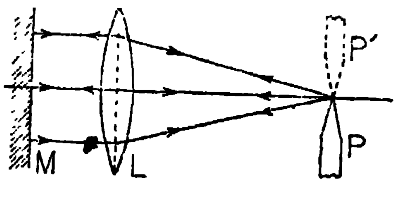


Fig. 217

Method. Place the lens L (Fig. 218) on a plane mirror M levelled previously. Mount a pin on an adjustable clamp. Looking from the top the pin and its image will be seen. When the pin is at the position P_2 you will see a real and inverted image at P'_2. Now adjust the pin so that its tip touches the tip of the image and there is no relative motion between them (as at $P_1P'_1$) when you move your eyes sideways. Measure the distance from the pin to the lens (LP_1). This is the focal length.

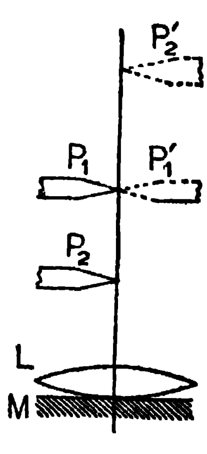


Fig. 218

Exercises

A. Spherical Mirrors

1. What is a spherical mirror? Explain the terms radius of curvature, principal axis, pole, principal section and focal length of a spherical mirror.

2. Deduce a simple relation between the focal length and the radius of curvature of a spherical mirror.

3. Prove the relation $\frac{1}{u}+\frac{1}{v}=\frac{1}{f}$ in the case of a concave mirror. Explain the notations used.

4. Deduce an expression for magnification in the case of a concave mirror.

5. Show graphically how a concave mirror forms a real image of an object.

How would you use a concave mirror to throw an image of a candle flame on the wall?

6. Where should an extended object be placed before a concave mirror so that an erect and magnified image is formed? Draw a diagram to show this. How is a shaving mirror used?

7. An object 1cm high is placed on the axis of a concave mirror of focal length 10 cm. If the distances from the mirror to the object be 5 cm and 20 cm respectively, find the nature, position and size of the image in each case. Do it graphically and verify by calculation. (*Ans*: Virtual, erect, 10cm, 2 cm ; real, inverted, 20 cm, 1 cm).

8. An object is placed 50 cm away from a spherical mirror. The image is formed at a distance of 75 cm on the same side of the mirror. What type of mirror is it and what is its focal length? (*Ans*: Concave; 30 cm).

9. A concave mirror of focal length 1 ft is at a distance of 5 ft from a wall. Where should a candle flame be placed to get a magnified image of it on the wall? How many times is it magnified? (*Ans*: 1¼ ft from the mirror; 4)

10. At what distance should an object be placed before a concave mirror of focal length f so that the image is (*i*) 1/4th the

size of the object, (*ii*) 4 times the size of the object? State the nature of the image in each case.

(*Ans*: $5f$; $\frac{5}{4}f$; Real and inverted).

B. Lenses

1. Why is a convex lens called a converging lens and a concave lens a diverging one? Illustrate with diagrams the action of a lens on a parallel beam of light.

2. Define first principal focus, second principal focus, focal length and the focal plane of a lens.

3. What properties of a lens may be utilized to find the position of the image of an extended object placed on the principal axis of a lens? Draw a diagram to illustrate your answer.

4. Draw a diagram showing the formation of an image of an extended object placed on the principal axis of a lens and hence deduce the lens formula.

5. What do you understand by magnification produced by a lens? How is the magnification related to object and image distances?

6. Draw diagrams to show how a convex lens can produce a magnified (*a*) real, (*b*) virtual, image?

A real image twice the size of the object is produced when an object is placed at a distance of 5 cm from a convex lens. What is the focal length of the lens? Where should the lens be placed so as to produce a virtual image of the same object magnified 3 times? (*Ans*: $3\frac{1}{3}$ cm; $\frac{20}{9}$ cm from the object.)

7. An object 1 in high is placed perpendicularly on the principal axis of a convex lens of focal length 10 in. If the successive distances of the object from the lens be 5 in and 20 in find the nature, position and size of the image in each case. Find it graphically and verify by calculation.

(*Ans*: Virtual, erect, 10 in. in front of the lens, 2 in; Real inverted, 10 in. behind the lens, $\frac{1}{2}$ in.)

8. An object is placed 6 cm in front of a lens, when an image is formed at a distance of 18 cm on the other side of the lens. What type of lens is it and what is its focal length?

(*Ans*: Convex ;$4\frac{1}{2}$ cm.)

9. An object 2 in high is placed at a distance of 6 in from a concave lens of focal length 15 in. What is the distance, size and nature of the image?

What would happen if the lens were convex and of the same focal length?

(*Ans*: 30/7 in, virtual and erect; 10 in;
$3\frac{1}{3}$ in, virtual and erect.)

10. What type of lens would you use to throw the image of a candle flame on a wall? What should be the focal length of a lens if an image magnified 10 times, is thrown on a wall by placing the candle flame at a distance of 10 cm from the lens?

(*Ans*: Convex; $9\frac{1}{11}$ cm.)

11. An object is placed at a certain distance in front of a convex lens of focal length 10 cm when the image is formed as far behind the lens as the object is in front of it. Find graphically the object distance and the magnification.

(*Ans*: 20 cm; 1.)

12. You have got two convex lenses. If you hold a pin in front of the first convex lens at a distance of 25 cm an inverted image is formed at a distance of 100 cm. If the pin is held in front of the second convex lens keeping the same object distance, the image distance also remain the same but the image becomes erect. Find the focal length of the two lenses.

(*Ans*:—20 cm; $33\frac{1}{3}$ cm.)

13. Define power of a lens. What is a dioptre. Two lenses of power $+2D$ and $-1{\cdot}5D$ are placed in contact. Find the power and the focal length of the combination.

(*Ans*: $\cdot 5D$;—200 cm.)

CHAPTER 25

OPTICAL INSTRUMENTS

In this chapter we shall give an account of some of the simpler optical instruments in daily use. Mirrors, prisms and lenses are used in instruments designed for various purposes. The action of the latter depends on the properties of the former.

200. The Photographic Camera

The essential parts of a photographic camera, as shown in fig. 219, are as follows:

(1) **A light tight chamber** whose interior is blackened. The side walls often consist of a folded leather bag which enables the distance between the front and the back of the camera to be varied.

(2) **A photographic objective** (*L*) (or simply, *the lens*), which consists of a number of lenses, the combination acting as a *single convergent* lens. It is placed in front of the camera, and is mounted on a slide so that it can be moved forwards and backwards. The purpose of the objective is to form a *real image* of the object to be photographed on a photographic plate.

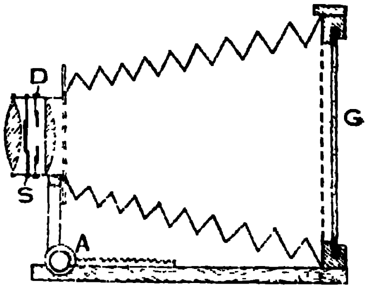


Fig. 219

(3) **A photographic plate** which is mounted at the back of the camera. It consists of a glass or a celluloid plate coated with an emulsion of silver halogen compounds in gelatin. When light falls on the emulsion silver is deposited in it, the thickness of the deposit depending on the intensity of light.

(4) **A shutter** (*S*) which is a device for shutting out or admitting light from the object into the camera.

There are generally quite a number of accessories in a camera of which the following may be mentioned:

(*i*) *A screen* (*G*), which consists of a ground-glass plate placed at the back of the camera. A sharp image of the object to be photographed is first formed on this screen. It is then replaced by the plate which is generally kept inside a plate-holder.

(*ii*) A *diaphragm* (*D*). It is an adjustable aperture which controls the amount of light received from the object. It lies generally inside the lens system forming the objective, or behind it.

Action of a camera. A sharp image of the object to be photographed is first formed on the ground-glass screen by moving the objective forwards or backwards as necessary.

The screen is then removed and replaced by the photographic plate. Light from the object is allowed to fall on the plate for a short duration depending on the intensity of the light and the relative aperture of the lens.

The *relative aperture* of a lens is defined as the ratio of the diameter of the aperture to the focal length of the lens. The larger the aperture, the shorter the exposure necessary. The f-number of a camera lens is the reciprocal of its relative aperture. A lens of f-number 4·5 means that the ratio of its aperture to the focal length is $1/4{\cdot}5$ or that the aperture is $f/4{\cdot}5$ where f is its focal length.

Light incident on the plate produces chemical reaction in the emulsion and deposits silver at different points of the image in different densities. This image is latent and needs to be developed and fixed by proper chemical treatment of the plate

The image so formed on the plate is a '*negative*', as the brighter portions of the object deposit more silver on the plate and appear darker. A *positive* is printed from the negative by placing a sensitized paper in contact with the negative and exposing the paper to light. This positive needs developing and fixing more or less like the negative.

201. The Eye

The eye is a camera in miniature. As an optical instrument it is far more perfect than any camera which human ingenuity has devised. It forms a real image on a sensitive screen called the *retina* and can focus objects at different distances.

Description. The **eyeball** (Fig. 220) is a nearly spherical chamber slightly compressed between the back and the front, and can be rotated in the socket by means of six muscles. The outer coating S, called the **sclerotic**, consists of a firm, fibrous, white tissue and serves to protect the eye from injuries. The front part of the sclerotic is transparent and is called the **cornea** (*C*). The sclerotic is coated on the inside with a dark coloured membrane, the **choroid** (*CH*) which prevents the reflection of scattered light and thus shields the retina from stray light. At the front of the eye the choroid merges into the **iris** (*I*) which has a hole called the **pupil** (*P*) at the centre. The iris is a *diaphragm* which, by involuntary muscle control, regulates the light admitted through the pupil. Behind the iris lies the **crystalline lens** (*L*) which is connected to the walls of the eye by the suspensory ligaments (*SL*). The back surface of the lens is more curved than the front surface and its refractive index increases inwards. Between the lens and the cornea is a watery liquid, the *aqueous humour*, while between the lens and the retina there is a jelly like substance, the *vitreous humour*. The line joining the centres of the cornea and the crystalline lens is called the **optic axis** of the eye.

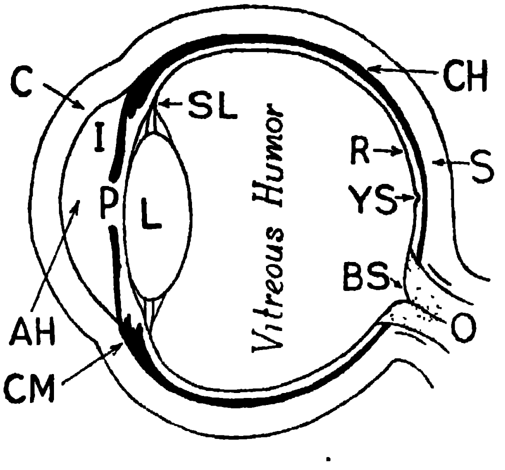


Fig. 220

The choroid merges on the inner side into an almost transparent rose coloured membrane called the retina (*R*). The retina is rich in nerve fibres and blood vessels, and is the part of the eye sensitive to light. All these nerves are joined directly to the brain, and where they leave the eye they form a bundle of nerves called the *optic nerve* (*O*). The region where the optic nerve penetrates the retina is not sensitive to light and is known as the blind spot (*BS*). It lies nearer to the nose than the yellow spot (*YS*) which is the most sensitive part of the retina. The image of an object directly viewed is formed on the yellow spot.

Action of the eye. The cornea, aqueous humour, (*AH*) crystalline lens and vitreous humour form together an optical system, and act like a single converging lens. For the normal eye, the focal length of the combination is such that distant objects are imaged on the retina. The process by which the eye adapts itself to see objects at different distances is called **accommodation**. By means of a ring-shaped muscle, known as the *ciliary muscle* (*CM*), attached to the crystalline lens the curvature of its surfaces (chiefly that of the front surface) may be increased, thus diminishing the focal length of the lens. The eye can thus see objects which are nearer.

The shortest distance at which an eye can see clearly is called the **near point**. Its distance from the eye depends on the age of the observer. Many persons, particularly children, can see clearly at a distance of 10 cm by exerting the maximum accommodation. This produces strain in the eyes. We are however accustomed to focus our eyes on objects at a distance of 20 to 30 cm. *Conventionally, the value* 25 *cm is taken as the least distance at which a normal* *eye usually sees clearly.* This distance is called the **least distance of distinct vision.**

A normal eye, called an *emmetropic* eye, can see distant objects without accommodation. The most distant point which can be seen without accommodation is called the **far point**. *The range of vision of a normal eye is therefore from the near point to infinity.* The eye focuses itself for different distances by the process of accommodation.

The eye forms on the retina *a real, inverted image* of objects which lie between its near point and far point. In spite of this inversion the brain interprets the images correctly (perhaps by experience).

202. Defects of Vision

The chief defects of vision which can be corrected by the use of lenses are the following:

(*i*) **Short-sight or Myopia.** If an eye causes parallel light to meet at a point in front of the retina it is said to be *myopic* or *short sighted* (Fig. 221). The far point is nearer than infinity, usually quite close to the eye. The near point may lie within a few centimetres of the eye. The defect is due to the focal length of the eye being too short or the eye ball being too long.

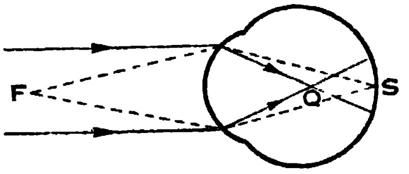


Fig. 221

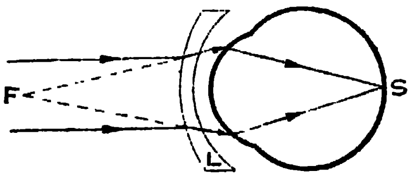


Fig. 222

The remedy is to use a concave spectacle lens of appropriate power which cancels the too great converging effect of the cornea and the eye lens, and enables images of distant objects to be formed on the retina.

In fig. 222 *F* is the far point of a myopic eye. If a concave lens with its focus at *F* be placed in front of the eye, distant objects will be clearly imaged on the retina. Otherwise the image will be formed at *Q*. (Fig. 221).

(*ii*) **Long-sight or Hypermetropia.** If the eye has too long a focal length or the eye-ball is too short, images of distant

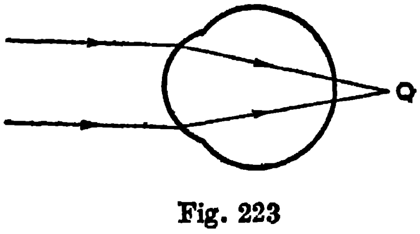


Fig. 223

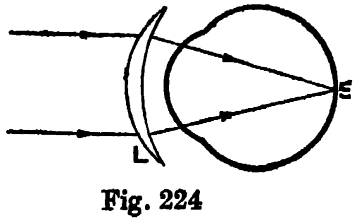


Fig. 224

objects will be formed behind the retina (Fig. 223). Such an eye is said to be *long-sighted* or *hypermetropic*. The defect can be corrected by using a convex lens of appropriate power (Fig. 224). The lens increases the convergence of the beam after refraction and enables the image to be formed on the retina.

(*iii*) **Presbyopia or Far-sight.** With age an eye may lose its power of accommodation and the near-point may recede beyond the normal 25 cm. Such an eye is said to be *far-sighted* or *presbyopic*. A person with this defect will hold a book at arm's length for reading.

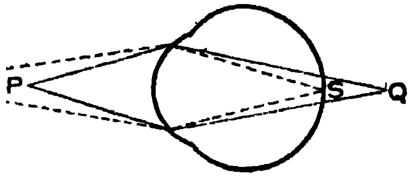


Fig. 225

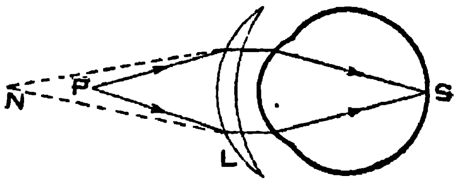


Fig. 226

The defect can be corrected by using a convex lens of appropriate power. In fig. 225, P represents the normal near point and N, the near-point of a presbyopic eye. The correcting lens L (Fig. 226) should have such a focal length that it can form at N a virtual image of an object placed at P.

203. Optical Aids to Vision and Angular Magnification

The apparent size of an object is determined by the size of the retinal image, which again is determined by the angle subtended at the eye by the extremities of an object. When an object is so small or so far away that details in it cannot be distinguished optical aids are used to enable the eye to view the objects with maximum comfort and sufficient detail. Instruments like microscopes and telescopes have been devised so as to increase the size of the retinal image by making the image subtend a larger angle at the eye than the object. The magnification, called the **subjective or angular magnification,** in such cases is the ratio

$$\frac{\text{angle subtended by the image at the eye.}}{\text{angle subtended by the object at the eye.}}$$

204. Magnifying Glass. The size of the retinal image of a small object can be increased and the object made to appear bigger by using a convex lens. A lens so used is called a **magnifying glass,** *hand-lens* or sometimes a **simple microscope**. There are two methods of using a hand-lens.

(*i*) **Image at the near point.** The angle subtended at the eye by a small object of height h placed at the near point, distant d from the eye, is h/d. The object should not be brought nearer to the eye than this as it would produce ill effects. Let us now place a

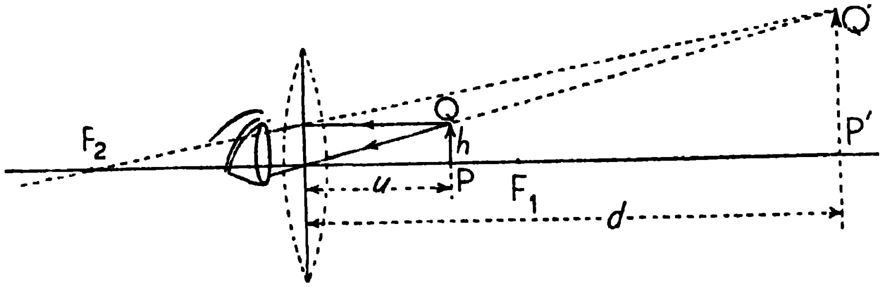


Fig. 227

convex lens in front of the eye and close to it, and move the object nearer to the lens so that a *virtual image is formed at the near point* (Fig. 227). Assuming the lens to be very close to the eye and ignoring the thickness of the lens, the angle subtended by the image at the eye is h/u.

Therefore the **subjective (or angular) magnification** $=\frac{h}{u}\Big/\frac{h}{d}=\frac{d}{u}$

But u is the object distance and d, the image distance. From the lens formula $1/v-1/u=1/f$, we get, on taking signs into consideration,

$$\frac{1}{d}-\frac{1}{u}=-\frac{1}{f} \tag{204.1}$$

Or, multiplying by d and transposing,

$$\frac{d}{u}=1+\frac{d}{f}$$

Hence the angular magnification

$$m=1+\frac{d}{f} \tag{204.2}$$

It follows, therefore, that a magnifying glass in conjunction with a *fully accommodated* eye gives a magnification which is greater the shorter the focal length of the lens. Such a glass is most useful for far-sighted persons, as d is larger. For short sighted persons d is small and little benefit is derived by using a lens as above. The image may be formed at any point between the near and the far points. The further away it is the smaller is the angle subtended at the eye, and hence the magnification is also smaller. Magnification also diminishes if the eye is moved away from the lens.

(ii) **Image at the far point.** The other method of using a hand-lens is to place the object in the focal plane of the lens so that the image is formed at the far point. This has the advantage that no accommodation need be exerted by the eye, and hence there will be no strain in the eye even if the work is continued for a long time. The angular magnification may be shown to be equal to d/f. The very small difference in magnification from the former case is not of much consequence, while it is a positive gain to avoid eye-strain.

205. The Compound Microscope

The purpose of the compound microscope is to increase the size of the retina[1] image of objects so small that the magnification produced by a hand-lens does not suffice.

Description. It consists of two convergent (convex) systems of lenses mounted coaxially in a tube. The lens near the object, called the *objective*, is of small aperture and short focal length. The other lens, behind which the eye is placed, is called the *eye-piece*. Its focal length is comparatively longer. The eye piece can be pushed in or drawn out as required.

Action. The action of a microscope will be clear from Fig. 228. Let L_1 be the objective and L_2 the eye piece. A small object PQ is placed just beyond the focus F_1 of L_1. Rays diverging from Q will, after refraction through the objective, converge to the point Q_1, forming a real, inverted and magnified image P_1Q_1. P_1Q_1 now serves as an object for the eye piece. The eye piece L_2 is so moved that P_1Q_1 falls within its focal distance. Rays diverging from Q_1 will after refraction through L_2 appear to come from Q', a virtual and magnified image of the object being finally formed at $P'Q'$. The eye piece functions just like a magnifying glass.

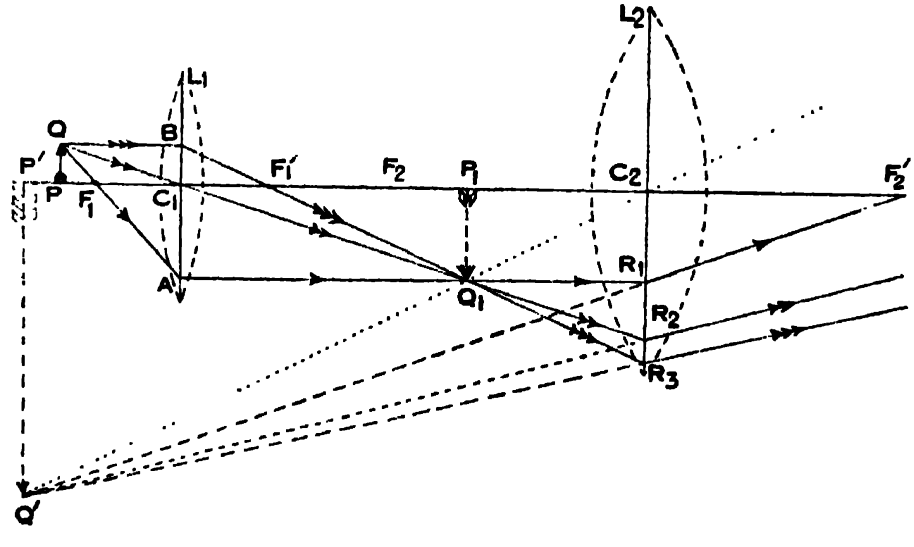


Fig. 228

The final image is usually formed at the near point though it may with profit be formed at the far point. The image at the near point gives maximum magnification, whereas at the far point it gives least eye strain.

Magnification. The magnification is produced in two stages : first by the objective and then by the eye piece. The resultant magnification can be shown to be equal to LD/f_of_e, where L=tube length ; D=least distance of distinct vision ; f_o, f_e=focal lengths of the objective and the eyepiece respectively.

To get a high magnification, f_o and f_e must be small. The magnification also increases with increase of tube length.

206. Telescopes

The purpose of a telescope is to increase the size of the retinal image of a distant object.

A real, inverted and diminished image of the distant object is first formed by a lens, called the **objective** of the telescope. This image is then viewed by a lens (actually a system of lenses), called the **eyepiece,** in the same way as a small object is seen through a magnifying glass.

A. Astronomical Telescope

Description. It consists of (1) an objective which is generally a combination of two lenses acting as a single convergent lens of *large focal length and aperture.* It forms a real, inverted and diminished image of a distant object in its focal plane.

(2) *A convergent eyepiece mounted co-axially with* the objective and having *its first focal plane coincident with the second focal plane of the objective.* The eyepiece thus acts as a magnifier to view the real image formed by the objective. *The final image is at infinity.* The distance between the objective and the eyepiece is the *sum of the focal lengths* of the two.

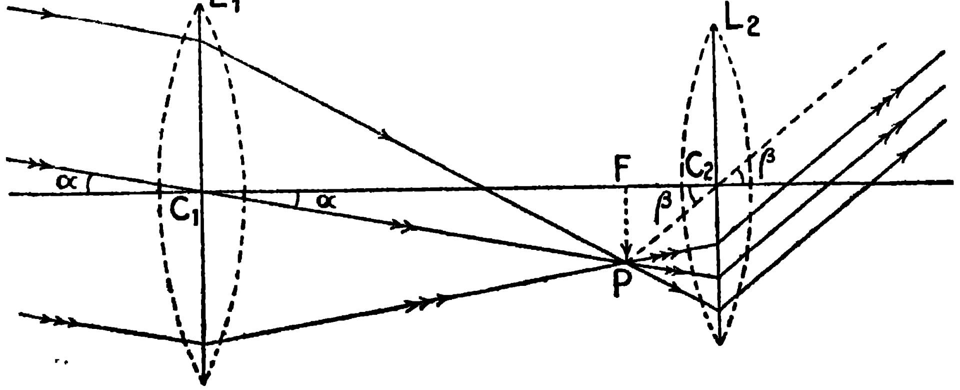


Fig. 229

Action. The optical action may be understood from Fig. 229. Since the object is situated at a great distance, the incident rays from a point on the object may be treated as a parallel pencil making a small angle with the axis of the telescope. Parallel rays from a distant object, on refraction through the objective L_1, form a real inverted and diminished image FP in the focal plane of the objective.

FP serves as the object for refraction by the eyepiece. For normal vision, the focal plane of the eyepiece is made to coincide with that of the objective. Thus FP becomes an object in the focal plane of the eyepiece L_2. So the rays diverging from FP, on refraction through L_2, are rendered parallel and a magnified and inverted image is produced at infinity. (The final image can be formed anywhere between the far and the near points of the observer by moving the eye piece towards FP. But having it at infinity minimizes eye-strain).

Magnification. The magnification m produced by a telescope is defined as the ratio

$$\frac{\text{angle } (\beta) \text{ subtended by the image at the eye}}{\text{angle } (\alpha) \text{ subtended by the object at the eye}}$$

Since the object is far away the angle subtended by it at the eye is practically the same as that subtended at the centre of the objective. Hence from Fig. 229.

$$m = \frac{\angle FC_2P}{\angle FC_1P} = \frac{FP/C_2F}{FP/C_1F} = \frac{C_1F}{C_2F} = \frac{f_1}{f_2} \qquad (206.1)$$

where f_1 and f_2 are the focal lengths of the objective and the eyepiece respectively.

An astronomical telescope is generally used for *viewing celestial bodies.* Since they are at great distances and appear very small a large magnifying power is needed to examine smaller bodies. Hence the focal length of the objective should be large. Moreover the brightness of the image depends on the amount of light from the object which the objective collects. Objectives of astronomical telescopes are therefore made to have a large aperture.

B. Galilean Telescope

The astronomical telescope is not suited for observation of terrestrial objects as the final image is inverted. For terrestrial observations a Galilean telescope may be used. It differs from an astronomical telescope in the nature and position of the eyepiece. The objective is similar to that of an astronomical telescope. The eyepiece is a concave lens. The distance between it and the objective is equal to the difference between the numerical values of their focal lengths. This makes the second principal focus of the objective and the first principal focus of the eyepiece coincident in position. The lenses are coaxial.

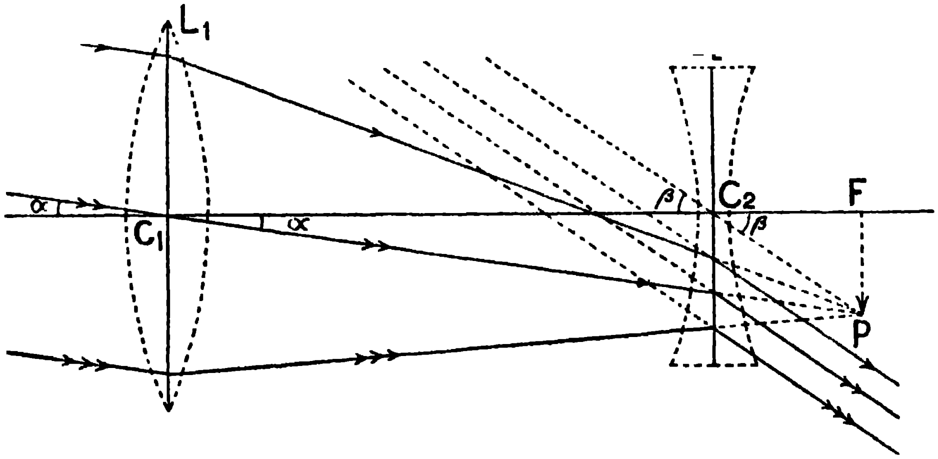


Fig. 230

In Fig. 230, the parallel pencil of ray from a distant object, on refraction through the objective L_1, would have formed a real image FP in the focal plane had not the concave lens L_2 intercepted it. Thus FP may be considered as a virtual object in the first principal focal plane of L_2. Consequently rays on refraction through the lens L_2 become a parallel beam. The final virtual image is thus at infinity and lies on the same side of the axis as the object. It is, therefore, erect.

Magnification. As before

$$m = \frac{\text{angle } (\beta) \text{ subtended by the image at the eye}}{\text{angle } (\alpha) \text{ subtended by the object at the eye}}$$

$$= \frac{\beta}{\alpha} = \frac{FP/C_2F}{FP/C_1F} = \frac{C_1F}{C_2F} = \frac{f_1}{f_2}$$

207. The Magic Lantern or Projection Lantern

The magic lantern is used for projecting upon a distant screen an enlarged image of a brilliantly illuminated transparent object, generally a photograph or drawing on glass. Its essential parts are :

(*i*) A powerful source of light (S; Fig. 231). A reflector (R) is placed behind the source to increase the illumination.

(*ii*) Two condensing lenses (C,C) of large aperture. The whole of the light they collect passes through the object.

(*iii*) A slide-holder (*AB*) containing the slide (*O*) which is the object.

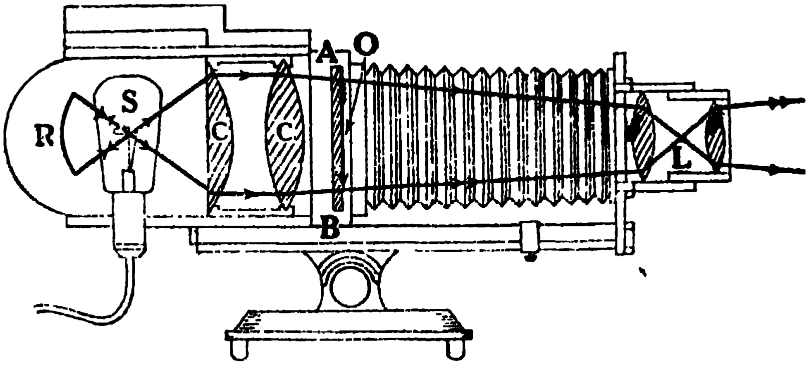


Fig. 231

(*iv*) An objective *L*, which is a combination of lenses forming a real, enlarged image of the slide on a screen.

Since the screen is at a considerable distance the slide is placed a little beyond the principal focus of the objective. Note that the lantern is much like a camera, but the positions of the object and image have been interchanged. We may call the lantern a camera in reverse. It is very important to have a powerful source of light which much be small in size

The condensing lenses form an image of the source at the centre of the objective. This increases the brightnecs of the inage to a maximum. The image is focussed on the screen by moving *L* as in a camera.

Exercises

1. Describe a photographic camera and explain how it works.

2. Draw a neat sketch of the human eye and label the parts. Describe its optical action. What are the similarities between a camera and an eye ?

3. What is accommodation ? What is meant by 'the near point' and 'the far point' of an eye ? What are the distances of these points in the case of a normal human eye ?

4. What are the main defects of vision ? How are they corrected ? Give diagrams.

5. How would you use a convex lens to act as a magnifier ? Where would you prefer to form the image and why ?

6. Describe a compound microscope and explain its optical action with a neat diagram.

On what factors does the magnifiying power of a microscope depend ?

7. Describe the construction and explain the optical action of an astronomical telescope.

On what factors does its magnification depend ? Why is the objective made large in size and focal length ?

8. Describe and explain the action of a Galilean telescope. How does it differ from an astronomical one ? What is its advantage ?

9. Draw a neat sketch of the projection lantern. Describe the function of the principal parts.

In what sense can it be called a camera in reverse ?

CHAPTER 26

DISPERSION AND SPECTRUM

208. White Light consists of Seven Colours

When a narrow beam of white light is incident on a prism, the emergent beam is not only deviated, but is at the same time split up into a coloured band. On placing a screen to receive this band, colours are found to lie in the order violet, indigo, blue, green, yellow, orange and red. The coloured band is called the **spectrum** of the incident light and the above sequence of colours is referred to as VIBGYOR. If any one of these colours is allowed to pass through a slit and fall on another prism, it will be further deviated, but no new colours will be formed. This shows that white light consists of seven colours. The colours in a spectrum gradually pass from one to the other. So the total number of colours is really infinite.

The splitting of white light into a coloured band occurs in a direction perpendicular to the refracting edge of the prism. Red is found to be least deviated and violet most. *The refractive index of red light is therefore smaller than that of violet light.* The separation of different colours due to the differences in their refractive indices is called **dispersion.**

209. Recomposition of White Light

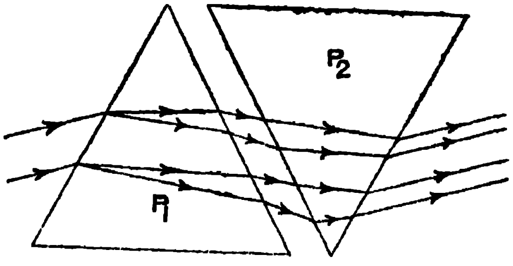


Fig. 232

(*i*) To verify whether the colours into which white light is split up by a prism may be combined to give white light again, the spectrum formed by a prism (P_1 ; Fig. 232) is allowed to pass through a second identical prism (P_2) turned the other way round. It will be observed that the second prism has combined the different colours into white light. The first prism has analysed the white light and the second prism has synthesised it.

(*ii*) Take a Newton's disc, which is a circular disc of card board divided into unequal sectors painted with the colours of the spectrum in their correct proportions (Fig. 233) and attached to a table. When the disc is rotated rapidly, it appears nearly white, not because there is any real mixture of colours, but because of the fact that impressions on the retina last for about $\frac{1}{10}$ second. Before the impression due to any one colour has died away it is succeeded by all other colours. This is a physiological, but not a physical mixture of colours.

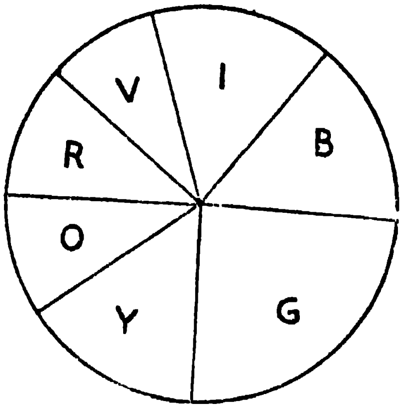


Fig. 233

210. Pure Spectrum

The spectrum obtained by allowing a beam of light to pass through a prism and receiving it on a screen is not pure. It consists of a series of coloured images of the source, one overlapping another (Fig. 234). This may be ascertained by blowing smoke in the path of the rays following the prism. The smoke appears coloured only at the boundary of the beam, the centre being white near the prism. Colours appear more clearly the greater the distance from the prism.

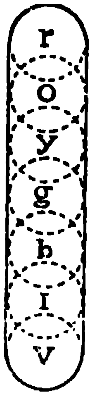


Fig. 2 4

A spectrum in which lights of different colours are not superposed is called a *pure spectrum.*

Conditions for producing pure spectra. The spectrum is purer the narrower the region over which the rays of a single colour from the source are concentrated. This may be brought about by

(*i*) using a narrow slit as the source of light ;

(*ii*) setting a prism for minimum deviation of the mean ray of the spectrum through it ;

(*iii*) Reducing the divergence or convergence of the bundle of rays (of a given colour) which passes through the prism. The best arrangement is to have a parallel beam.

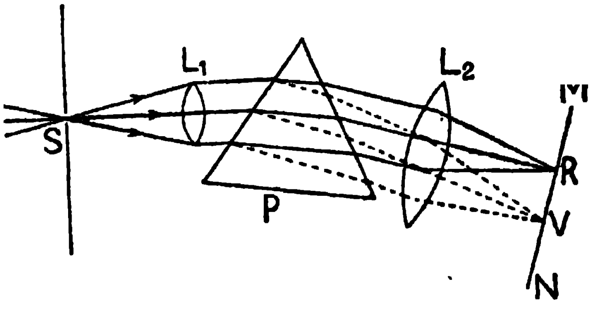


Fig. 235

A method of producing pure spectrum. The above conditions are satisfied in the arrangement represented in fig. 235. *S* is a narrow slit illuminated by white light. It is in the focal plane of the lens L_1. L_1 forms a parallel beam, which fall on the prism *P* interposed behind L_1 in the position of minimum deviation for rays of the middle part of the spectrum, i.e., for yellow-green rays. Emergent rays of any one colour undergo the same deviation and come out as a parallel bundle. Owing to dispersion the angle of emergence is different for different colours. A second convex lens L_2 now brings the different parallel bundles to different foci, red at R, violet at V, with other colours in between them. The spectrum RV can be received on a screen placed in the second focal plane of L_2.

211. Types of Spectra

A. Emission Spectra. The spectrum of the light emitted by an incandescent body is called its *emission spectrum.* Emission spectra may be classified as (*a*) *continuous spectra,* (*b*) *line spectra* and (*c*) *band spectra.*

(a) Continuous spectra. Emission spectra in which all colours are found to be present are known as continuous spectra. Such spectra are emitted by all *incandescent solids,* and are more or less alike.

(b) Line spectra. Spectra consisting of narrow discrete regions only (the so-called 'lines') are called line spectra. They are emitted by *atoms* of elements in the gaseous state. Each element has its characteristic spectrum; it can, therefore, be identified by its spectrum. The number of lines vary widely from element to element (Fig. 236). While sodium has practically one line in the visible spectum, iron has thousands.

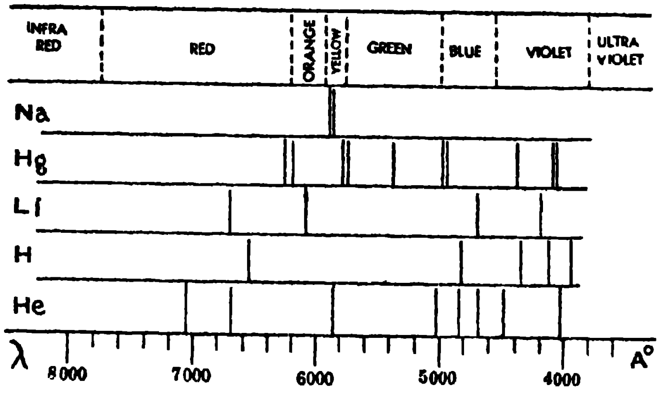


Fig. 236

(c) Band spectra. Spectra which appear to consist of 'bands' of light are called band spectra. Each band consists of a large number of lines grouped together. Band spectra are emitted by *molecules* in the gaseous state.

B. Absorption spectra. When white light is passed through a coloured, transparent substance and the transmitted light is spectrally analysed, it is found that some regions of the continuous spectrum are missing. These missing regions constitute an absorption spectrum.

Each substance has a definite absorption spectrum and can be identified with its help. A man was acquitted of murder charge when it was found by spectrum analysis of the blood stains on his cloth that the blood was that of a goat and not of a man.

Solar spectrum and Fraunhofer lines. The spectrum of the sun, at first sight, appears to be a continuous one. But on closer examination it is found tobe crossed by a large number of fine dark lines. It is actually a *line absorption spectrum*. The fine dark lines are called **Fraunhofer lines** after the name of the discoverer. Fig. 237 represents the solar spectrum in which some of the principal Fraunhofer lines have been shown. They are named by the letters A, B, C, D etc, and are due to absorption by atoms of various elements, such as H, He, Ca, Na, etc. Over fifty elements known on earth have been identified in the sun by their spectrum.

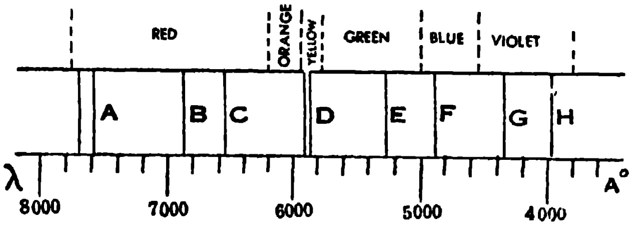


Fig. 237

212. Colours of Bodies

The colour of a body depends on (*i*) the light which illuminates it and (*ii*) the light it reflects if it is opaque, or the light it transmits if it is transparent.

If a red rose is illuminated by light of different colours, as in the spectrum of sun light, it will appear red only in red light and practically black in light of other colours. The rose looks red because it reflects red from the light falling on it and absorbs the other colours. If the light illuminating it has no red in it, there is nothing for the rose to reflect; hence it looks black. If an assortment of flowers of different colours is illuminated by light from a sodium flame, which is yellow, white and yellow flowers will appear yellow, and the rest dark.

A white body looks white in day light because it reflects equally all the colours of day light falling on it and does not absorb any. A green leaf appears green in day light because it reflects green and absorbs the other colours of day light. In red light the white body will look red and the green leaf, dark. We thus see that *the colour of an opaque body depends on the colour of the light it reflects.*

If a spectrum is viewed through a piece of red glass, only the red portion can be seen. A green glass will transmit only the green.

When the red glass is held in green light or the green glass in red light, it looks almost black. When we hold the red and green glasses together in day light the combination appears dark as almost no light passes. These show that *the colour of a transparent body depends on the colour of the light it transmits.*

Mixture of coloured lights. Almost any colour can be produced by a suitable combination of *three* colours. They are *red*, *green* and *blue*, and are called the **primary colours**. Most colours can be matched by superposing on a screen illuminations of these colours in different proportions.

Two colours which together produce the sensation of white are called **complementary colours**. If from the spectrum of white light we take away red, the rest of the colours appears bluish green. Red and this bluegreen, which is sometimes called peacock green, are therefore complementary colours. Yellow and indigo, green and purple or magenta, orange and blue, are similar pairs of complementary colours.

Mixture of pigments. When yellow and blue light are superposed on a screen the resulting colour is nearly white. But when a yellow pigment and a blue pigment are mixed in equal proportions the colour of the resulting mixture is *green*. This is so because yellow pigment reflects both yellow and green, and the blue pigment reflects both blue and green, each absorbing all colours of the spectrum other than those it reflects. As green is the only colour which both reflect the mixture of the two pigments appears green. *The colour of a mixture of pigments is the one which escapes absorption by the different ingredients.*

Exercises

1. How would you show that white light is made up of many colours, and that these colours may be combined to produce white light ?

2. What is a pure spectrum ? How would you form one ? Draw a neat sketch to represent your arrangement.

3. How would you classify the different kinds of spectra? What are the kinds of spectra to which gaseous atoms, gaseous molecules and incandescent solids give rise?

4. What kind of spectrum is the solar spectrum? What are Fraunhofer lines?

5. Describe simple experiments to show that the colour of a body depends on

(*a*) the colour of the light which illuminates it,
(*b*) the light it reflects, if it is opaque,
(*c*) the light it transmits, if it is transparent.

6. What are meant by *primary colours* and *complementary colours*?

When a yellow and a blue light are superposed on a screen the colour is nearly white; but when a yellow and a blue pigment are mixed the colour is nearly green. Explain.

CHAPTER 27

GENERAL PROPERTIES OF MAGNETS

213. Natural and Artificial Magnets

Modern civilization owes a great deal to magnets. Large scale production of electrical power would have been impossible without them. We would have no electric motors driving machinery in our factories, or running our tramcars and electric trains. No communication through telegraphy and telephony would have been possible. Examples of this kind may be multiplied without limit.

Knowledge of natural magnets dates back to about 3000 years. The Chinese are reported to have known even about 1000 B.C. that small bits of iron would stick to a certain kind of stone, later called lodestone. It is an ore of iron known as magnetite and occurs in the earth's crust at many places, such as Asia minor, Scandinavia, the U.S.A., and some other countries.

The name *lodestone* was given to those specimens of magnetite which had the following two properties:

(*i*) **An attractive property.** A piece of this stone attracts iron. The power of attraction is mostly concentrated at two places of the stone.

(*ii*) **A directive property.** When a piece of this stone is suspended horizontally it tends to come to rest with these ends pointing roughly north and south.

It is easy to procure specimens of lodestone; almost any dealer of scientific instruments will be able to supply one. To test the attractive property the stone may be dipped into iron filings. The filings will adhere to two ends of the stone. Small iron nails will be similarly attracted (Fig. 238).

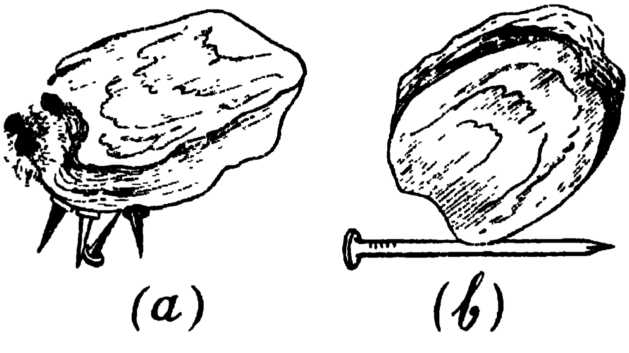


Fig. 238

To test the directive property the specimen may be placed on a paper stirrup and hung by a long thread from a wooden stand (Fig. 239). The thread unwinds at first causing the stone to spin. Bring the spin to a stop and see how the stone takes up a fixed direction. If disturbed it will again take up this direction—very nearly a north-south one. It is as if some external controlling force acted on the stone and repeatedly brought it to rest in the same position. *The controlling force is due to the magnetism of the earth.* In ancient times mariners are said to have guided their ships with the help of the directive property of these stones. Hence the name lodestone or lead-the-way stone.

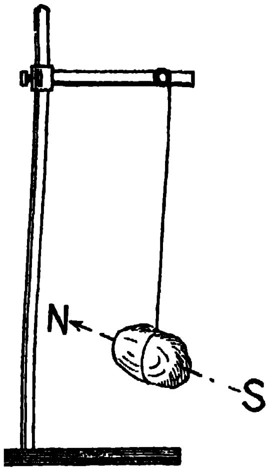


Fig. 239

Bodies which have the attractive and directive properties described above are called **magnets**. Lodestones are *natural magnets.*

It soon came to be discovered that these properties could be imparted to small pieces of iron by rubbing them with a lodestone Pieces of iron which have been made to acquire the attractive and directive properties of natural magnets are called *artificial magnets.* They can be made much more powerful than natural magnets and are the kind we use practically for all experimental purposes. When a body is endowed with the attractive and directive properties of a natural magnet, it is said to be *magnetized* or to have acquired the property of *magnetism.*

Artificial magnets are generally named according to their shape. Ordinarily we have the following (Fig. 240):

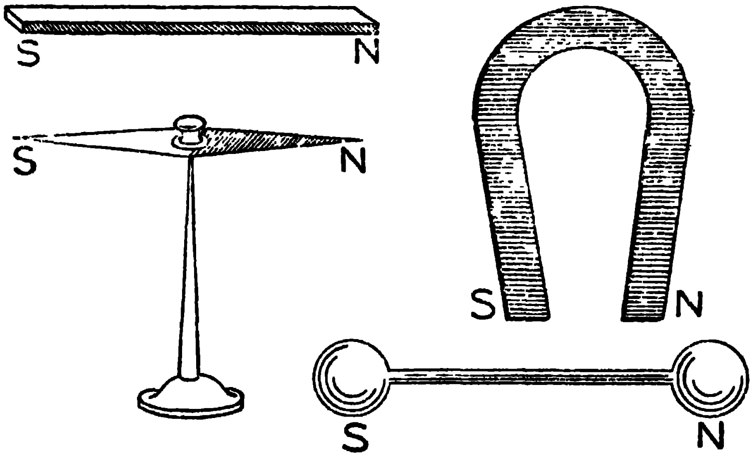


Fig. 240

1. *Bar magnet.* It is a bar of rectangular or circular section.
2. *Magnetic needle.* It may have any of the shapes shown in fig. 241. The magnet is so pivoted as to be horizontal and free to move in a horizontal plane. It is a most useful piece of apparatus and may be small or large (Also see § 245). Small needles encased in a glass case with or without the cardinal points indicated are

generally called *compass needles.* Some needles may have a long, light pointer fixed to it at right angles.

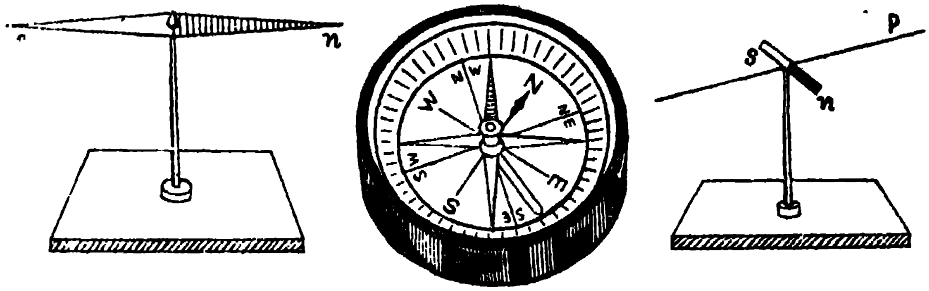


Fig. 241

3. *Horse-shoe magnet.*
4. *Ball-ended magnet.*

The list is by no means exhaustive.

214. Some Definitions

(i) **Pole.** When a bar magnet is dipped into iron filings, they adhere mostly to the ends of the bar (Fig. 242). The attractive property of a magnet is strongest near the ends. The *poles* of a magnet are the regions where the attracting power is greatest. A magnet has two poles at its two ends. The middle of the magnet has practically no attractive power ; iron filings hardly stick to this region.

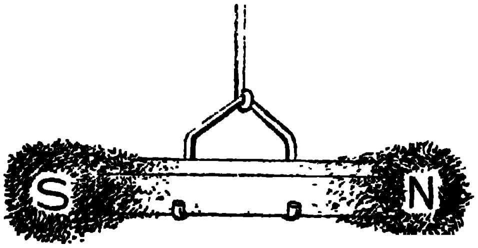


Fig. 242

(ii) **The North Pole and the South Pole.** When a magnet is suspended freely in a horizontal plane (Fig. 243), it will be seen to come to rest with a particular end pointing north. It will take the same position even when displaced. The pole near the end of the magnet which points north is called the **north-seeking pole** or the **north pole.** The other pole, which is towards the south, is called the **south-seeking pole** or the **south pole.**

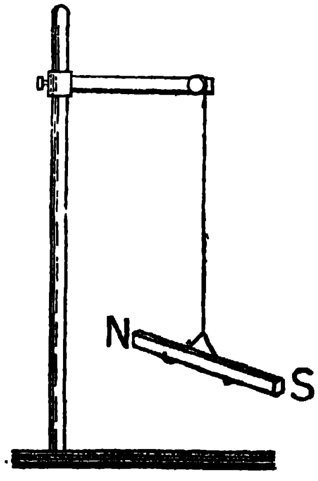


Fig. 243

(iii) **Magnetic Axis.** The straight line joining the two poles of a magnet is called the magnetic axis. It does not always coincide with the geometrical axis.

(iv) **Magnetic length.** The distance between the poles of a magnet is called the magnetic length. It is shorter than the geometrical length (Fig. 244).

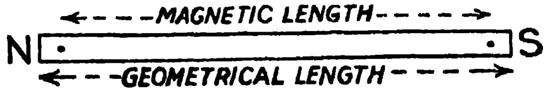


Fig. 244

(v) **Magnetic meridian.** When a freely suspended magnet comes to rest, the vertical plane passing through the axis of the magnet is called the magnetic meridian of the earth at the place.

215. Like Poles repel ; Unlike Poles attract

Take a magnetic needle or suspend a bar magnet so that it can move freely in a horizontal plane. Take another magnet of which the poles are marked. From a considerable distance bring the north pole of the latter near the north pole of the former along a line perpendicular to the former (Fig. 245). You will find that there is repulsion between the two.

Repeat your observation bringing the south pole of one near the south pole of the other. There will be repulsion. *Like poles repel.*

Now bring the north pole of one near the south pole of the other and vice versa. There will be attraction. *Unlike poles attract.*

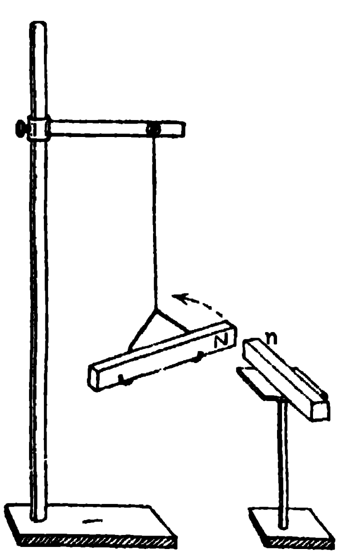


Fig. 245

216. Repulsion is a surer Test of Magnetization

If a bar of unmagnetized iron is brought near a pole of a magnet there will be attraction between the two as between two unlike poles. Hence when there is attraction between two bars one of which is known to be a magnet, we cannot say definitely whether the other is also a magnet or not. If, however, there is repulsion between two ends, we can say definitely that both must have the same polarity, and hence the other bar must also be a magnet. Repulsion is therefore a surer test of magnetization than attraction.

217. How to test a Piece of Material for Magnetization

To test if a body is magnetized we should, from a distance, bring one end of it *gradually* close to one end of a magnetic needle along a line perpendicular to the needle. (Also see §223). If there is repulsion, the body is magnetized. That end of it which approached the needle had the same polarity as the pole it repelled.

If there was attraction, the observation should be repeated on the other pole of the needle. If there is attraction again, the body is unmagnetized. If we find repulsion now, we should conclude that the body is magnetized.

218. Magnetic and Non-magnetic Substances

Substances which are strongly attracted by magnets are generally called *magnetic* substances. They are few in number and ordinarily comprise iron, nickel, cobalt and some of their alloys. Present day high quality magnets are made of various alloys, mainly of the elements iron, nickel and cobalt. Alnico (an alloy of composition 10 Al, 18Ni, 12 Co, 54 Fe, 6 Cu) is one of the best materials for making permanent magnets. It is hard and brittle, and cannot be machined satisfactorily. Magnets made of alnico alloy therefore have a crude look. Ticonal and cobalt steel (with 35% Co) also make strong permanent magnets.

The great majority of substances do not appear to be affected by magnets and are commonly referred to as being *non-magnetic*. This classification of substances into magnetic and non-magnetic is rather crude as magnets have some action on all substances, though extremely feeble in most cases.

It may be instructive to discuss how to identify three exactly similar looking bars one of which is a magnet, one a bar of unmagnetized iron (a magnetic substance) and the third, a bar of brass (a non-magnetic substance).

In doing so we may remember that

(*i*) The brass bar will neither attract the other bars nor be attracted by them as brass is non-magnetic;

(*ii*) The bar of unmagnetized iron will be attracted at all places over it by either end of the magnet;

(*iii*) the middle of the magnet will show no attractive power.

So, if we take the bars in turn and run one end of the bar so taken along the other two, one after another, we should conclude as follows:

(*a*) If the bar in hand in not attracted by the other two, it is the barass bar.

(*b*) If the bar in hand is attracted by the ends of another bar, but not by the middle portion of the latter, the one in hand is the unmagnetized iron bar, and the other, the magnet.

219. Magnetic Induction

Place a magnetic needle on the table (Fig. 246). Along a line perpendicular to the axis of needle and passing through one of its ends (poles), place a magnet *NS* at the same height. Let the N-pole of *NS* point towards the N-pole of the needle. Adjust the distance between the needle and the magnet such that the repulsive action between their north poles is hardly perceptible.

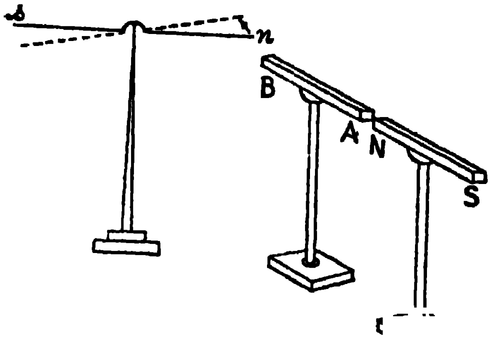


Fig. 246

Now place a *soft iron bar AB* between the magnet and the N-pole of the needle. The latter will be found to be repelled. This shows that bar *AB* has acquired magnetism under the influence of the magnet *NS*, and that the end *B* has north polarity. Obviously *A* has a south polarity. On reversing the magnet *NS* so that its south pole is nearest to *A*, it will be found that the N-pole of the needle is attracted by *B*. The poles on *AB* have been reversed by reversing the poles of *NS*.

When the magnet *NS* is removed, the bar *AB* will be found to have practically lost all magnetism. Magnetism on *AB* reappears as soon as *NS* is put back in its position.

It thus appears that the bar *AB* acquires magnetism only when it is under the influence of the magnet *NS*. This sort of magnetism is called *induced magnetism* and the phenomenon is called **magnetic induction.**

The N-pole of *NS* induced a S-pole at the end of *AB* nearest to it. **An inducing pole induces opposite polarity at that end of the body under induction which is nearest to it, and similar polarity at the remote end.**

The soft iron bar *AB* may be replaced by similar bars made of other materials to see if they are capable of being magnetized by induction. Repeating the experiment with various materials it will be seen that magnetism can be induced only in magnetic substances, some, such as soft iron, being more easily magnetizable than others.

220. Induction precedes Attraction

It is reasonable to think that when a magnet attracts iron, events take place in the following sequence :

(*i*) As soon as one pole of a magnet is brought near a piece of iron, its end nearest to the magnet acquires by induction a polarity opposite to that of the attracting pole, while the farthest end acquires the same polarity.

(*ii*) An attraction then ensues between the attracting pole and the induced opposite pole.

There is also a repulsion between the inducing pole and the induced like pole. But the opposite poles being nearer, the attraction is greater than the repulsion.

The attraction between a magnet and a piece of iron (or any other magnetic material) is thus a result of magnetic induction. By the statement "induction precedes attraction" we mean the above sequence of events.

221. Additive Effect of Induction

Place a bar of soft iron on a table with about a third of it projecting beyond the edge. Keep two strong magnets on the table with their like poles close to one end of the bar (Fig. 247). They will magnetize the bar by induction. At the other end support as many small iron nails (or gramophone pins) as possible. Gradually move away one of the magnets. You will find that some of the nails fall off. Reverse the magnet removed and bring its opposite pole gradually closer to the bar. You will find that more nails fall off and finally the bar may hold none.

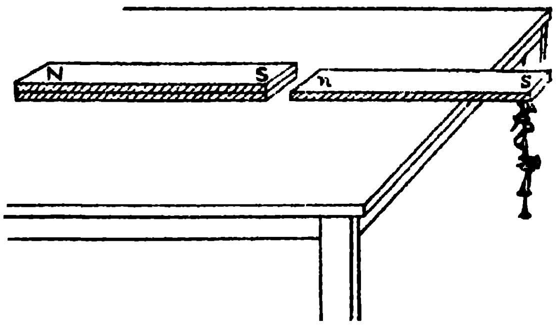


Fig. 247

This shows that the strength of the induced magnetism depends on the strength of the inducing pole. The former increases or diminishes as the latter. *Induction is an additive effect of the inducing poles.*

222. Amount of Induced Magnetism

The amount of magnetism induced in a body depends on the following factors :

(*i*) The strength of the inducing pole. The stronger it is the greater is the magnetism induced.

(*ii*) The distance between the inducing pole and the induced body. The shorter this distance the greater the induction.

(*iii*) The nature of the magnetic substance. Magnetic induction in a soft iron piece will be greater than in a similar piece of steel under identical conditions. That in cobalt or nickel will be still less.

223. Change of Polarity due to Induction

When the north pole of a strong magnet is brought sufficiently close to the north pole of a magnetic needle (or a weaker magnet), an attraction is often seen to occur between the two. The reason is that the strong N-pole induces on the N-pole of the needle a south polarity which is stronger than the existing north pole of the needle. The stronger induced pole overcomes the effect of the weaker permanent pole. Hence the pole behaves like a south pole and is attracted by the inducing N-pole.

Such change of polarity may be permanent. This fact must be remembered while testing the polarity of a magnet with the help of another magnet. The two magnets should be brought *gradually nearer together from a distance.* A strong pole should never be brought too close to a testing pole.

224. Induction through Non-magnetic and Magnetic Substances : Magnetic Screening

Induction takes place without hindrance even if non-magnetic materials intervene ; but it is effectively hindered by the intervention of magnetic materials. This may be tested as follows :

Support a strong magnet vertically on a wooden stand (Fig. 248). Tie a soft iron nail to the stand by a short piece of string so that the nail is held in air by the attraction of the magnet, but a little away from the attracting pole. Adjust the distance between the two so that any further increase in distance causes the nail to drop.

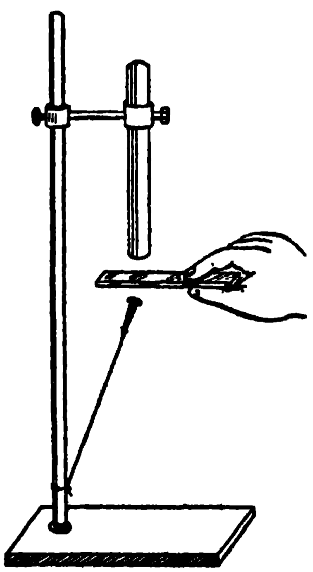


Fig. 248

Slide sheets of non-magnetic material between the nail and the magnet. The nail is unaffected, showing that the inducing effect of the magnet on the nail is not reduced by the interposition of a non-magnetic material between the two. Induction takes place unhindered through non-magnetic materials.

If a sheet of magnetic material is introduced between the nail and the magnet, the former immediately drops. This shows that the effect of the magnet on the nail is reduced by the interposition of a magnetic material between the two.

Magnetic substances differ in their power of so reducing the effect of a magnet. It has been found that substances which are most easily magnetized are the most effective. Soft iron is better than steel in this regard, and steel better than cobalt or nickel.

If, therefore, we want to keep a space free from the effect of magnets, we have to surround the space with sheets of soft iron. This is known as *magnetic screening.*

225. Some important Magnetic Properties : Susceptibility, Retentivity, Coercivity

(a) **Susceptibility.** When we say a boy is susceptible to cold, we mean he catches cold easily. In magnetism the term 'susceptible' has a qualitatively similar meaning. A substance which acquires induced magnetism more easily than another is said to have a higher susceptibility.

Soft iron has a higher susceptibility than steel. In the presence of an inducing magnet soft iron acquires a stronger magnetism than a piece of steel. There is an alloy called *mu-metal* which can be very strongly magnetized by even a weak magnetizing force. It has a high susceptibility. Its composition is 73 Ni, 22 Fe, 5 Cu.

(b) **Retentivity.** It is that property of a magnetic substance by virtue of which it retains magnetism after the inducing magnet has been removed. The magnetism retained is called *residual magnetism* or *remanence.*

Steel retains more than soft iron the magnetism it acquires under given conditions. Hence we say that steel has a higher retentivity than soft iron. Permanent magnets should be made of materials with high retentivity.

(c) **Coercivity.** Substances with a residual magnetism behave differently when attempt is made to remove it. Some substances resist the removal of the residual magnetism more strongly than others. Coercivity is that property of a magnetic substance by virtue of which it retains magnetism even when subjected to forces tending to reduce the magnetism.

A permanent magnet should be made of a material with a high coercivity as well as high retentivity. Alnico (§ 218) is such a material. Two alnico magnets, placed one on top of the other, repel so strongly that the top magnet may remain suspended in air if prevented by 'guides' from twisting or shooting to one side.

226. Methods of Magnetization

A magnetic substance may be magnetized in a number of ways.

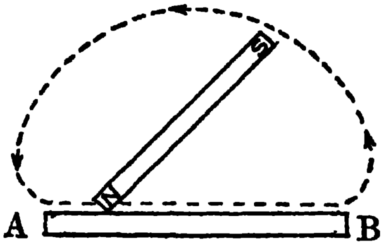


Fig. 249

(i) **Method of single touch.** Fig. 249 shows the method. *AB* is the bar to be magnetized, and *NS* the magnet with which to do it. *AB* is stroked repeatedly with one pole of *NS*, starting always at one end and finishing at the other. *NS* must be lifted and brought back to the starting point, as shown by the dotted line in the figure. *AB* is then turned upside down and the strokes repeated.

The end of the bar where the stroking magnet leaves it, acquires a polarity opposite to that of the stroking pole.

The method is suitable only for magnetizing thin bars such as compass needles. Darning and knitting needles can also be magnetized in this way. *NS* should be a strong magnet. The effect clearly is one of induction.

(ii) **Method of divided or separate touch.** The method is shown in fig. 250. The stroking is done with two magnets, opposite poles being used over the two halves in the same way as in the previous method. The strokes are repeated with the iron bar turned upside down. The law of polarity is the same as in method (*i*). Instead of two magnets, the two poles of the same magnet can be used for stroking the two halves, starting always at the middle.

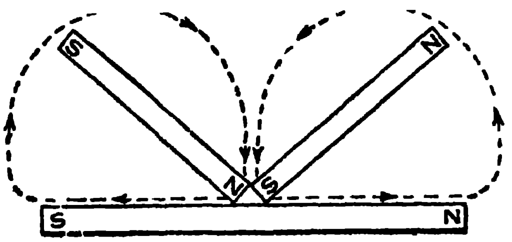


Fig. 250

This method gives stronger magnets than method (*i*).

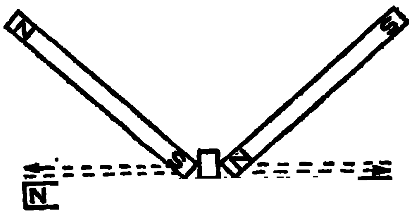


Fig. 251

(iii) **Method of double touch.** Here two stroking magnets are used. They are held obliquely with their opposite poles near each other but separated by a piece of cork or wood (Fig. 251). The oblique magnets are moved together from the centre of the bar to be magnetized to one end, then to the other, and finally brought to a stop at the centre. The operation is repeated a number of times, and also with the bar turned upside down. The polarities will be as shown.

This method produces stronger magnets than either method (*i*) or (*ii*).

None of the above methods are, however, of any practical use. Much stronger magnets can be produced by electrical means as described below. Besides, electrically produced magnets are free from one defect common to magnets produced by stroking. The latter often have like poles at the ends, and the opposite poles at the middle. Such poles are called *consequent poles.*

(iv) **Magnetization by electric current.**

(a) **Magnetizing solenoid.** A piece of glass tubing (Fig. 252) wound with close turns of insulated copper wire serves to magnetize knitting needles or rods or bars of iron and steel. The bar is placed inside the tube, and a strong electric current sent through the copper coil, called a *solenoid.* To produce better results the bar may be gently tapped with a wooden mallet while the current is passing. The current need not pass for more than a short while. On withdrawing the bar from the solenoid it will be found to have been magnetized. This method produces strong magnets free from consequent poles.

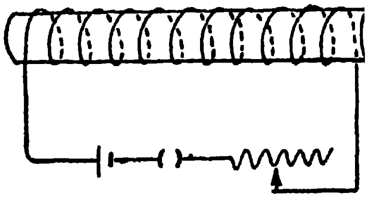


Fig. 252

Viewed from one end the current in the solenoid will appear to flow in a *clockwise* direction. The end of the bar lying towards this end of the solenoid will have *south* polarity.

(b) **Electromagnets.** If we take a soft iron rod and pass an electric current round it, as in the above experiment, the soft iron becomes a strong magnet. When the current ceases to flow the soft iron loses its magnetism immediately. A magnet the magnetism of which is excited by an electric current and lasts so long as the current lasts, is called an *electromagnet.*

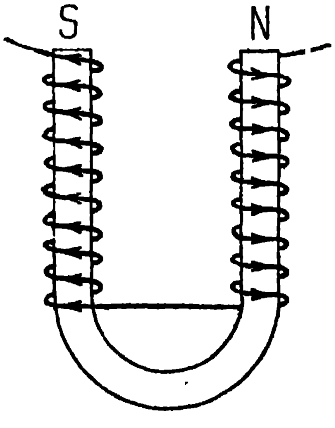


Fig. 253

Electromagnets have various shapes and are of many designs. But whatever the shape or design, an electromagnet will have its *magnetizing coil* (or coils) placed round a *core,* generally of soft iron. In most, the core has the shape of a U (Fig. 253). If, looking at one end of the coil or core, the current appears to flow in a clockwise direction, this end will have south polarity. If the direction is anti-clockwise, the end will have north polarity. The windings of the wire on the two limbs of a U-shaped core must therefore, be in opposite directions as shown in fig. 253.

Electromagnets are used for a large variety of purposes. Apart from their uses in the laboratory they are widely used in industry. They form an essential part of all electrical generators and motors. Relays used for transmitting messages over long distances by telegraphy and telephony, consist essentially of electromagnets. An electromagnet often used in factories for lifting large masses of iron and removing them from one place to another is shown in fig. 254. It is known as a 'pot' magnet. One pole is in the centre and the other is in the form of a ring surrounding it. When non-magnetic material has to be lifted, it is fastened by the ring to the lower iron plate. The plate is then lifted by the magnet. Such magnets may be several feet in diameter and can lift tons of iron.

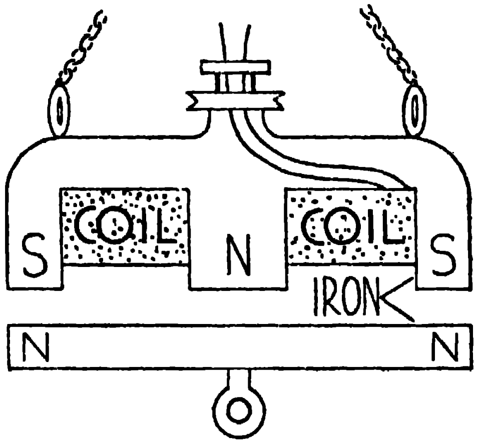


Fig. 254

A "cracker ball" electromagnet has a large iron ball weighing several hundredweights. It is picked up by the electromagnet mounted on a crane and released from a height to break and pulverize articles underneath. Electromagnets are also used for sorting scrap iron from other materials. In the surgical field it is used for extracting iron or steel splinters from the eye or other parts of the body.

A piece of magnetic material placed between the poles of an electromagnet will be magnetized more easily and strongly than by stroking it with permanent magnets. Magnetization is helped by lightly tapping the material when it is under the influence of the electromagnet.

227. Magnetization due to the Earth

We shall see later (Chap. 29) that the earth itself is a huge magnet. It can produce magnetization ; but the effect is very feeble and can be demonstrated only if materials of high magnetic susceptibility, such as soft iron, soft steel or special alloys like permalloy (80 Fe, 20 Ni) or mu-metal (73 Ni, 22 Fe, 5 Cu), are used. If a piece of the material is held horizontally in the magnetic meridian or vertically, and gently tapped it will acquire a magnetism. A better effect is produced when the specimen is placed in the magnetic meridian at a particular angle to the horizontal (the 'dip' angle; see § 242). In Calcutta this angle is 31°, the north end pointing below the horizon.

The north end acquires north polarity. When magnetized vertically, the lower end will have north polarity in the northern hemisphere of the earth.

Iron beams pointing north and south, and vertical iron railings are generally found to be slightly magnetized. Ships made of steel plates acquire a magnetism due to the earth during hammering and riveting. During World War II, Germans placed 'magnetic' mines in shallow waters. The feeble magnetism of a ship was enough to explode the mine by setting off a magnetic trigger in the mine. The fight against magnetic mines was won by surrounding a ship with a cable. The cable carried an electric current enough to neutralize the magnetic effect of the ship at a distance.

228. Laminated Magnets

We have seen in § 224 that the influence of a magnet is greatly reduced in passing through a magnetic substance. Hence when a thick bar is magnetized, the deeper lying layers of material will be much less affected than layers near the surface. If the material is in the form of a thin sheet, the whole of the material can be evenly magnetized. Strong permanent magnets can thus be made of a number of thin sheets of steel, each fully magnetized. The sheets are bolted together with their similar poles in the same direction (Fig. 255). Such magnets are called *laminated magnets*.

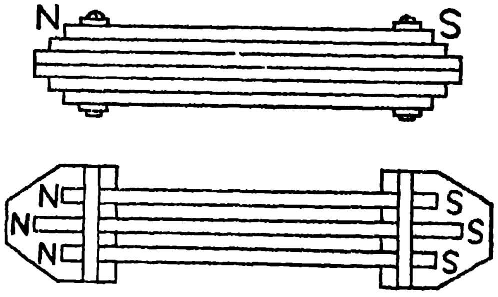


Fig. 255

229. Demagnetization

Demagnetization means the complete or partial removal of magnetism from a magnetized specimen.

A. Partial demagnetization can be brought about in a number of ways, such as by rough handling and by proximity of similar poles. Besides, a magnet exerts a demagnetizing effect on itself.

(*a*) *Rough handling*. When a magnet is subjected to mechanical shocks by dropping, hammering, twisting etc., it loses a part of its magnetism. Soft iron suffers more than steel under the same conditions. Materials with high coercivity, e.g. cobalt steel, alnico, ticonal etc. are less affected by rough handling.

(*b*) *Proximity of similar poles.* When two similar poles are brought near each other, one reduces the strength of the other by induction. If one of the materials has a low coercivity (e.g., soft iron or soft steel) the change in it may be permanent.

(*c*) *Demagnetizing effect of a magnet on itself.* A magnet with free poles exerts a demagnetizing effect on itself. The effect is due to induction and may be understood by considering a bar magnet. The north pole tends by induction to produce in the material of the magnet a south polarity towards itself and a north polarity away from it. This has the effect of weakening the south pole. The south pole also acts in a similar way and tends to reduce the strength of the north pole.

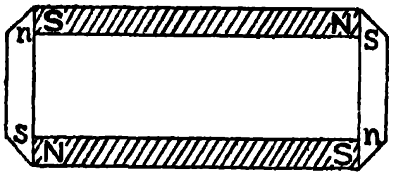


Fig. 256

To minimize this demagnetizing effect, bar magnets, when not in use, are kept parallel to one another with their opposite poles on the same side and connected by soft iron pieces, called **keepers** (Fig. 256). Each pole of the magnet induces an equal and opposite pole in the material of the keeper in contact with it. The effects of these two poles on any part of the magnet are equal and opposite. Horse-shoe magnets, when not in use, are also fitted with keepers for the same reason (Fig. 257). The demagnetizing effect is due to free poles, and the effect of keepers is to prevent free poles from appearing.

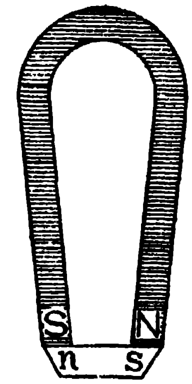


Fig. 257

B. Complete demagnetization can be achieved either by heating or by electrical means.

(*a*) *Heating.* When a magnetized material is heated beyond a certain temperature, which is characteristic of the material, it loses its magnetism completely. This characteristic temperature is called its **Curie point**. For iron and its alloys the Curie points lie in the range 750°C to 900°C. For nickel it is about 360°C.

Heating below Curie point causes a slight fall in the magnetism. On cooling the original strength is regained.

(*b*) *Demagnetization by electrical means.* The specimen to be demagnetized may be placed in a solenoid fed with an alternating current. It is then withdrawn without switching off the current.

Hair springs of watches are made of steel. If a watch happens to be brought close to a strong magnet, the hairspring is magnetized.

The watch then ceases to give correct time. The hairspring can be demagnetized by holding the watch near a coil supplied with and fed by a alternating current and gradually moving it away.

230. It is not possible to isolate a Pole

One might think that if a magnet were broken into two, each part will have one pole only. But it is not so. Take a piece of magnetized watch spring about 4 inches long and break it into two. You will find that each piece is a complete magnet and has opposite polarities at its ends. If the two new poles at the broken ends be dipped together into iron filings, they would hardly attract any. This shows that the new poles are opposite in sign, but equal in strength.

Further divisions produce no new result. Each piece, however small, turns out to be a complete magnet with two poles (Fig. 258).

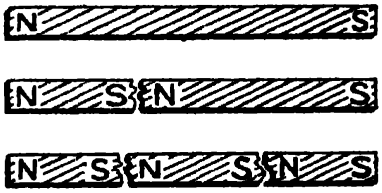
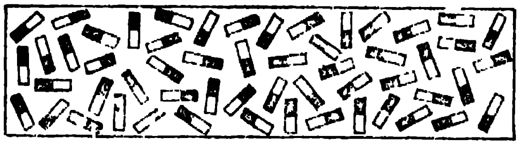


Fig. 258

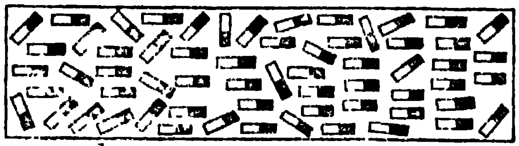
We may conceive these divisions and subdivisions of a magnet to be continued till we reach the molecular stage. Molecules of a magnetic substance should, in such a case, be looked upon as magnets.

231. Molecular Theory of Magnetism

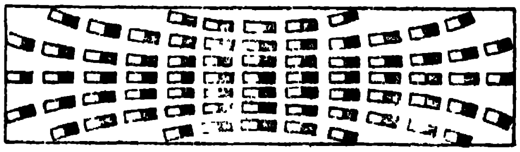
Many phenomena in magnetism can be explained if we assume that molecules of a magnetic substance are themselves magnets. You will perhaps immediately ask why every piece of iron does not behave as a magnet if it is made up of innumerable tiny molecular magnets. The reason is that in an unmagnetized specimen the molecular magnets are arranged in a haphazard manner, and even in a small region there are as many north poles pointing in a given direction as south. At all points outside the iron, the effect of the north poles will be neutralized by that of the south poles, and no resultant magnetism will appear. Fig. 259(*a*) indicates such an arrangement, the little



(*a*) UNMAGNETISED



(*b*) PARTLY MAGNETISED



(*c*) FULLY MAGNETISED

Fig. 259

rectangles representing a molecular magnet with its shaded and unshaded portions standing for the two polarities.

If a magnetizing force is applied to the material, the tiny magnets align themselves according to the direction of this force. Complete alignment does not come about unless the magnetizing force is strong enough. No further increase of magnetization is possible when all the magnets have been fully aligned. The specimen is then said to have reached *magnetic saturation.*

Partial and complete magnetization are depicted in figures 259(*b*) and 259(*c*) respectively.

Magnetization thus gives rise to long chains of molecular magnets. At one end of a bar there are only north poles, and at the other, only south poles. At the middle the two kinds of poles are adjacent to each other. When a magnet is broken at the middle, there will be a row of north poles on one side, and a row of south poles on the other.

In support of what we have said we may perform the following simple experiment. Take a test-tube about half full of iron filings, close it with a stopper and lay it on side on a table. Test with a compass needle if it has a polarity at any end. Ordinarily it will have none. Now stroke the tube with a permanent magnet as in magnetizing. Testing again with the compass needle you will find that it has become magnetized; the iron filings have partially oriented themselves under the action of the magnet.

Shake the tube and test again. You will find that all signs of magnetization have disappeared.

232. Magnetic Properties of Soft Iron and Steel

It has been necessary for us to mention the magnetic properties of soft iron and steel from time to time. Though there are many varieties of soft iron and steel, their magnetic properties may be broadly indicated as follows :

Substance	*Susceptibility or Ease of magnetization*	*Retentivity or ability to retain magnetism*	*Coercivity or difficulty in removing magnetism*
Soft iron	High	Low	Low
Steel	Low (More difficult to magnetize)	High	High

These properties make soft iron the material of choice for constructing the core of an electromagnet. Steel is the proper material from which to construct permanent magnets.

Special alloys, such as permalloy, mu-metal, have very high susceptibilities, and very low retentivity and coercivity. At the other end there are special steels and alloys, such as cobalt steel, alnico, ticonal, which have high retentivity and coercivity.

Exercises

1. What are the fundamental properties of a magnet ? Define the terms : *magnetic pole*, *magnetic axis* and *magnetic meridian.*

2. Why is repulsion considered a surer test of magnetization than attraction ? How would you test a specimen for magnetization ?

3. You are given three identical looking bars one of which is a magnet, one is made of soft iron and the third, of brass. How would you identify them without any other aid ?

4. What are the principal facts about magnetic induction ? How would you demonstrate them ?

5(*a*). The north pole of a magnetic needle may be attracted by the north pole of a strong bar magnet. How is this possible ?

(*b*). What is meant by the statement that induction precedes attraction.

6(*a*). State briefly what you understand by susceptibility, retentivity and coercivity. Discuss these properties of soft iron and steel.

(*b*). What kind of magnetic properties should a material have in order that it may be used for the construction of (*i*) a permanent magnet, (*ii*) an electromagnet ?

7. You are given a steel needle. What method would you adopt to magnetize it as best as possible ?

8. What is an electromagnet ? What determines its polarity ? Mention some of its uses.

9. What are laminated magnets ? What is their advantage ?

10. How can a magnet be demagnetized completely ? What is the function of *keepers* as are attached to horse-shoe magnets when not in use ?

11. Why is magnetism considered to be of molecular origin ?

12. Write short notes on (*i*) magnetic screening, (*ii*) Curie point (*iii*) consequent pole.

MAGNETIC FIELD OF FORCE

233. Magnetic Field

We have seen that a magnet can exert forces at a distance. *A magnetic field is a region over which magnetic forces may be felt.* Every magnet produces around it a magnetic field, which is ordinarily not perceptible beyond a metre or two. The magnetic field of an electromagnet may be detected even at a distance of several metres. The earth's magnetic field is perceptible miles above its surface.

The question naturally arises as to how to describe a magnetic field. It is not difficult to imagine that if we could place a free, isolated magnetic pole in a magnetic field, it would move under the action of the force it experiences. Though we cannot isolate a pole it is possible to find what the effect of a magnetic field on an isolated pole would be. Let us perform the following experiment.

Take a large and deep trough made of a non-magnetic material. A wooden vessel coated with wax and held together by wooden pins or brass nails will do. Place a strong permanent magnet at its centre at a height of about eight inches from the bottom on a non-magnetic stand. Fill the trough with water up to the level of the magnet. Take a knitting needle about six inches long and magnetize it. Float the needle vertically in water on a piece of cork with the north pole projecting a little out of the water. Bring the needle near the north pole of the permanent magnet and release it (Fig. 260). You will find that the needle moves away in a curved path towards the south pole. Release it at some other point; it will follow a different path. Such paths in a magnetic field are known as *lines of force.*

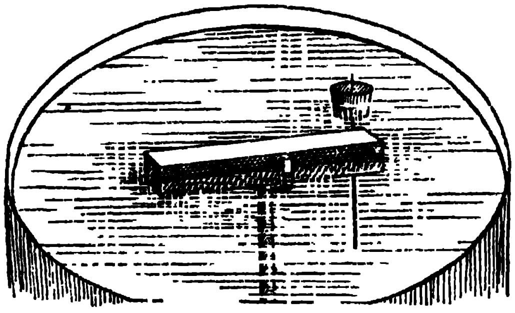


Fig. 260

234. Lines of Force

A line of force in a magnetic field is defined as the path that would be followed by a north magnetic pole if it were free to move in the *field.* The direction of the field at any point is taken as the direction of motion of the north pole as it moves along the line of force through the point. Since the direction of motion at any point of a path is tangential to the path, we may also define a line of force by saying that :

A line of force in a magnetic field is such a line that the tangent to it at any point gives the direction of the magnetic field at that point. We shall see later (§244) that the two poles of a magnet are equally strong. While the N-pole is urged along a line of force in one direction, the S-pole is urged with an equal force in the opposite direction. The result is that *a tiny magnet sets itself along the line of force without moving bodily in any one direction.* This gives a method of plotting magnetic fields (See § 235).

The concept of lines of force plays a very important role in magnetism. To explain magnetic phenomena they may be conceived to have the following *properties* :

(*i*) A line of force defines the path along which a free north pole would move if placed in the field.

(*ii*) The tangent to a line of force at any point gives the direction of the magnetic field at that point.

(*iii*) Lines of force originate on a north pole and end on a south pole. (This is so because the path is taken to be the one followed by a *north* pole. It will start by repulsion from a north pole and end by attraction at a south pole.)

(*iv*) Through every point in a magnetic field there is only one line of force. No two lines of force can intersect each other. (The force experienced by a pole in a magnetic field has a definite direction at each point of the field. If two lines of force intersected, there would be two directions for the field at that point. This is contrary to experience.)

(*v*) Each line of force has a tendency to contract longitudinally. (This will explain the attraction between a north and a south pole.)

(*vi*) A line of force tends to repel another laterally. (This will explain the repulsion between two like poles.)

235. Tracing Lines of Force in a Magnetic Field

There are two methods of tracing lines of force, one using iron filings, and the other using a small compass needle. In both, we utilize the property that a tiny magnet sets itself along the line of force.

Experiment 46. *To trace the lines of force due to a bar magnet with iron filings.*

Apparatus : Paraffined paper, iron filings, a sheet of glass, strong bar magnet.

Place the glass sheet on the magnet and the paraffined paper on the glass. Lightly dust iron filings from a muslin bag moving the bag all over the paper at a height of about six inches. Tap the glass sheet gently. The filings will arrange themselves on the paper in continuous curves. These curves (Fig. 261) describe the pattern of the lines of force. It can be fixed by slightly heating the paper. (The pattern can be formed more easily on glass; but it cannot be fixed.)

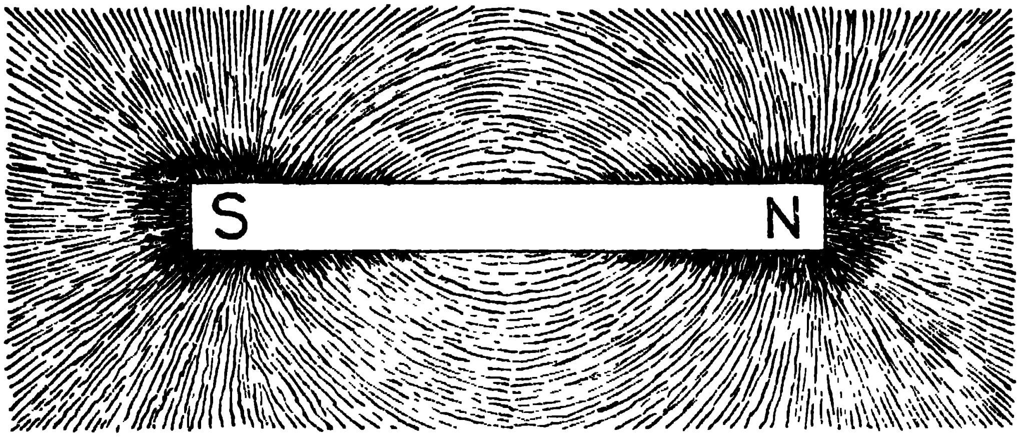


Fig. 261

Experiment 47. *To trace the lines of force due to a bar magnet with the help of a compass needle.*

Apparatus : Drawing paper, drawing board, small compass needle, strong bar magnet.

Fix the paper on the drawing board and place the bar magnet at its middle. Draw the boundary of the magnet and see that it does not move away from this position during the experiment.

Place the compass needle near one corner of the north pole of the bar magnet. Make a dot on the paper with a pencil as close to each pole of the compass as possible. Move the compass until its south pole is closest to the dot that previously marked its north pole (Fig. 262). Now make another dot close to its north pole. Move the compass again and keep repeating this process until the

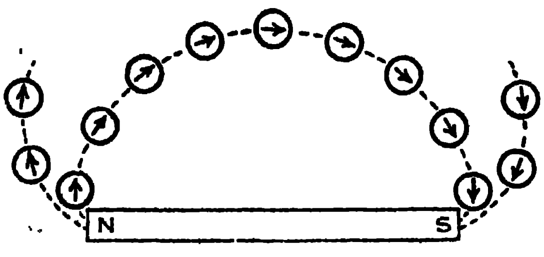


Fig. 262

compass has reached the south pole of the magnet. Connect all the dots with a smooth line. This represents a line of force.

The whole field may be mapped out in this way by changing the initial position of the needle.

Some patterns are shown below (Figs. 263 to 266). They are self-explanatory.

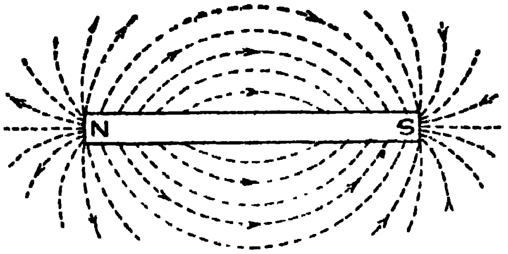


Fig- 263

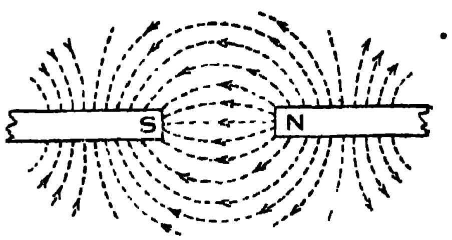


Fig- 264

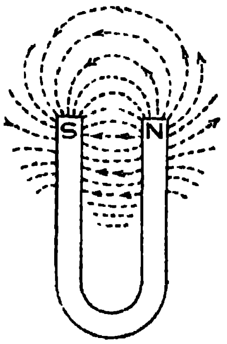


Fig. 265

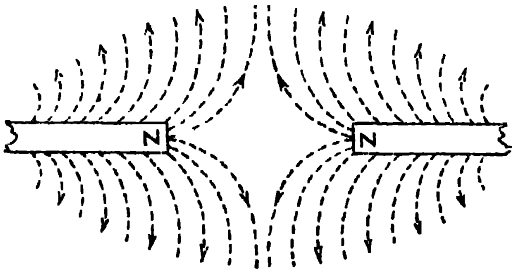


Fig. 266

236. Coulomb's Law

Some magnets attract more iron filings than others do. This shows that magnetic poles differ in their ability to exert a force. The quantity 'pole strength' or 'strength of a pole' is a measure of this ability.

From a series of carefully designed experiments Coulomb found that *the force between two magnetic poles varies directly as the product of their strengths and inversely as the square of the distance between them.* This is known as *Coulomb's law.*

It was also found that though a pole acted on another at a distance, the force between them was affected by the nature of the intervening medium. If m, m' denote the strengths of the poles and r the distance between them, the force between them is given by

$$F = \frac{1}{\mu}\frac{mm'}{r^2}$$

where μ is a constant depending on the nature of the medium between the poles, and is called the *permeability* of the medium.

For vacuum μ is taken as unity. In the case of magnetic substances the value of μ is high, while for all non-magnetic substances it can be taken practically to be equal to unity.

The Unit Magnetic Pole. Coulomb's law leads to the definition of the unit of magnetic pole strength. If two identical poles, placed 1cm apart in vacuum, repel each other with a force of one dyne, each is said to have unit strength and is called a *unit pole* (or a pole of unit strength). There is no special name for the unit of pole strength.

237. Intensity of a Magnetic Field

The *intensity* of a magnetic field at a point is defined as the force in dynes that acts upon a unit north pole when placed at the point in question. The terms 'magnetic field strength' and 'magnetic intensity' are synonymous with 'intensity of a magnetic field'.

The unit of intensity of a magnetic field is *one dyne per unit pole.* It is called the *Oersted.* (Oersted was a Danish physicist who, in 1820, discovered that an electric current produces a magnetic field around it.) Until recently it was called the 'gauss', a term which still persists in some books. (Gauss was a German mathematician and physicist.)

It follows from the definition of field strength that a magnetic pole of strength m units when placed in a magnetic field of intensity H oersteds will experience a force of f dynes given by

$$f=mH$$

238. Moment of a Magnet

The most important physical quantity associated with a magnet is its *magnetic moment.* It is numerically given by the product of the strength of either pole and the distance between them. If m is the pole strength and l the distance between them, the magnetic moment M is given by

$$M=ml.$$

It is possible to determine the moment of a magnet without knowing the pole strength and the distance between the poles separately.

Let a magnet of moment M be placed in a field of strength H. If the magnet is free to turn it will align itself in the direction of the field. If held forcibly at right angles to the field, the magnet will be acted on by a couple which will tend to align it with the field. The magnitude of the moment of this couple may be shown to be MH. Hence we may also say that the *moment of a magnet* *is equal to the mechanical couple required to keep the magnet at right angles to a field of unit intensity.*

*** One of the meanings of the word 'moment' is importance, as you will see from expressions like 'a matter of much moment, of no moment etc.' The term 'moment of a force' represents the effectiveness of a force in producing rotation. It has, therefore, been appropriate to give the name 'magnetic moment' to that property of a magnet which is of the greatest importance in describing its behaviour.**

Exercises

1. What are magnetic lines of force ? What are their fundamental properties ? How does a short magnet set itself relative to a line of force ?

2. State Coulomb's law of force between magnetic poles.

What is meant by permeability of a medium ? Does a magnetic substance have a higher permeability than a non-magnetic one ?

3. Define unit pole and intensity of magnetic field. What is the unit in which the latter is expressed ?

4. Give two definitions of magnetic moment.

CHAPTER 29

TERRESTRIAL MAGNETISM

239. The Earth behaves as a huge Magnet

The fact that a suspended magnet orients itself in a particular direction at every point on or near the earth, shows that the earth is surrounded by a magnetic field. Besides, pieces of iron are found to be magnetized, though slightly, when they lie in a north-south or vertical direction. This magnetization is increased by light mechanical shocks. A bar made of an alloy called mu-metal, when placed north-and-south, will show north polarity at its north end. If reversed in position, the new north end will show north polarity. Such behaviour is of the kind one meets in magnetization under the influence of a magnet, and could be explained by assuming that the earth itself is a huge magnet.

240. Simulation of the Earth's Field

The nature of the earth's field is such as might be produced by a huge bar magnet placed at the centre of the earth with its axis inclined at an angle of about 18° to the axis of the earth (Fig. 267). In the class room it may be simulated by a strong bar magnet, preferably an electromagnet, mounted inside a globe of non-magnetic material (Fig. 268). A glass tube may be fitted through the centre of the globe at an angle of about 18° to the axis, and the magnet slipped into it. The length of the magnet may be about half the diameter of the globe. A soft iron rod wound round with several layers of insulated copper wire may serve as the magnet. To excite it current may be passed through the coil from a dry cell. A small magnetic compass, as is often attached to a watch chain, may be placed at various points of the globe to find the

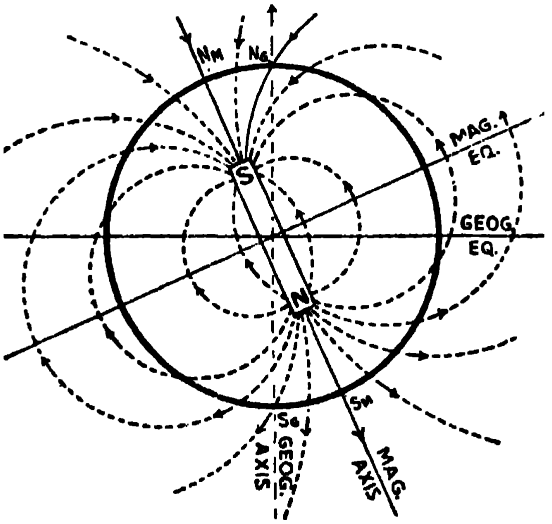


Fig. 267

direction of the magnetic field. The small magnets *ns* in the figure are intended to show such directions.

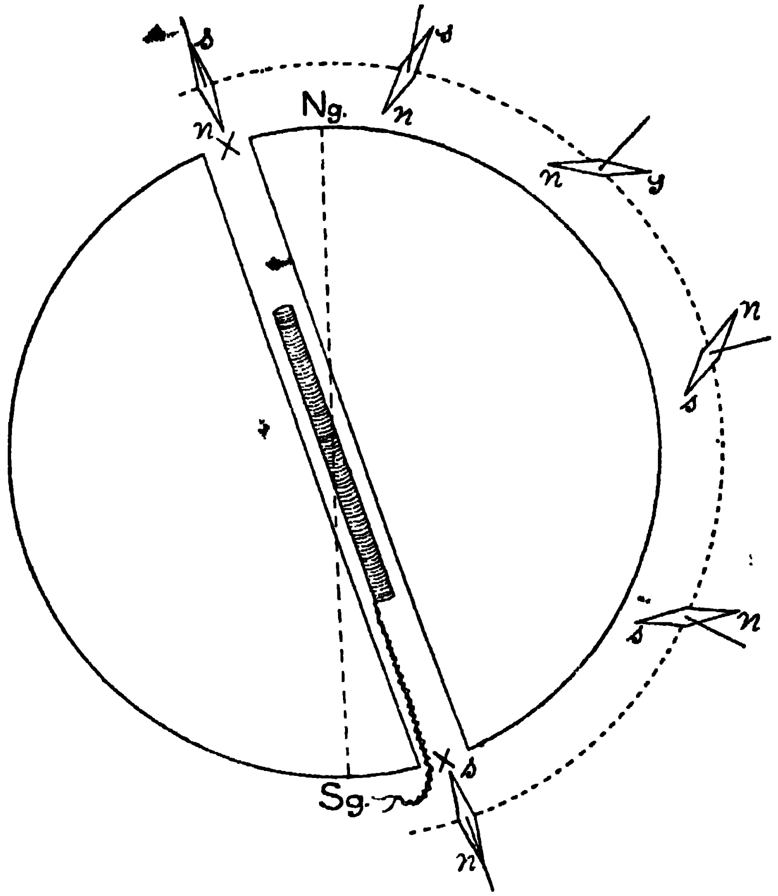


Fig- 268

241. Magnetic Poles of the Earth

A freely suspended magnetic needle will set itself vertically at two places on the surface of the earth. These places are known as the magnetic poles of the earth. One of them is near the geographic north pole and is called the magnetic north pole. The other is the magnetic south pole. Their present locations are approximately as follows :

N-pole : 74°N latitude ; 100°W longitude (It is in Northern Canada).

S-pole : 70°S latitude ; 150°E longitude. (The place is in Antarctica).

The world-wide scientific investigations carried out during the International Geophysical Year—July 1957 to December 1958—are expected to throw much light on the earth's magnetism, the origin of which is not yet known.

242. Magnetic Elements of the Earth's Field

To specify completely the earth's magnetic field at a place it is necessary to state (*i*) the *strength* of the field and (*ii*) its *direction*. The direction is that of the axis of a magnetic needle freely supported at its centre of gravity. A plane passing through this axis and the centre of the earth is called the *magnetic meridian*.

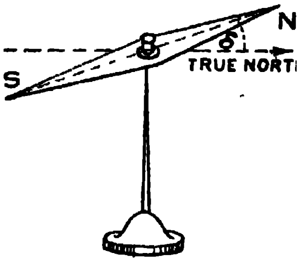


Fig. 269

The angle between the magnetic meridian and the true geographical north is called the *declination* (Fig. 269). It is also known as *deviation* or *variation*.

The angle between the direction of the earth's magnetic field and the horizontal, measured in the plane of the magnetic meridian, is called the **dip** or *inclination*.

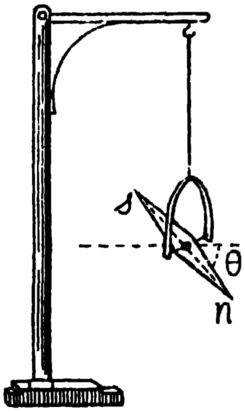


Fig. 270

Fig. 270 represents a magnetic needle freely supported at its centre of gravity. It rests in the magnetic meridian with its axis in the direction of the earth's field. The angle between the axis of the magnet and the horizontal in the plane of the magnetic meridian is the dip (θ). Dip is taken as positive when the north pole of the needle points downwards. It is so in the north hemisphere of the earth.

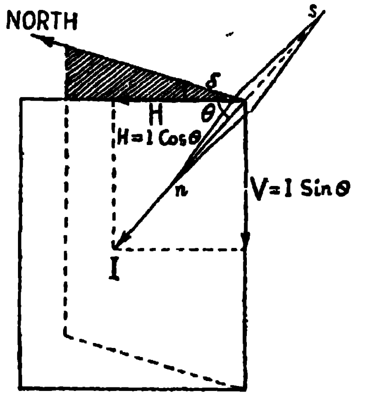


Fig. 271

In fig. 271 the shaded plane represents a vertical plane passing through the geographical north and south (the true meridian). The unshaded plane is the magnetic meridian. The angle δ between them is the declination. The needle *ns*, freely supported at its centre of gravity, points along the earth's field*. θ is the dip.

The total intensity or strength of the earth's magnetic field may be resolved horizontally and vertically in the plane of the magnetic meridian. The component of the intensity of the earth's magnetic field resolved along the horizontal in the plane of the magnetic meridian is called the **horizontal component of the earth's magnetic**

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* Note that an unmagnetized body, similarly supported, will have no tendency to turn and will remain at rest in any position.

field. Denoting this component by H, the vertical component by V, and the total intensity by I, we have

$$H = I\cos\theta$$
$$V = I\sin\theta$$
$$\tan\theta = \frac{V}{H}$$
$$I = \sqrt{H^2 + V^2}.$$

It is clear that the earth's field at a place will be fully specified if we have stated the values of (*i*) the dip (θ), (*ii*) the declination (δ) and (*iii*) the horizontal component (H) at the place. These three quantities are called the **magnetic elements of the earth.**

Table. Values of H, V, θ and δ at some places.

Place	H	V	θ	δ	*Place*	H	V	θ	δ
	(*in oersteds*)					(*in oersteds*)			
Calcutta	·37	·22	31°	40′E	New York	·18	·57	72°	10°W
Bombay	·37	·17	25°	32′E	Chicago	·18	·59	73°	2½°W
London	·18	·43	67°	10°W	Magnetic poles	0	—	90°	—

243. Measurement of Dip

The dip at a place can be measured with an instrument called *dip circle*. It has a magnetic needle (N, Fig. 272) capable of rotation about a horizontal axis passing through its centre of gravity. The needle is mounted in the centre of a vertical circle (A) divided into four quadrants each of which is graduated in degrees from 0 to 90. The 0°—0° line is horizontal and the 90°—90° line is vertical.

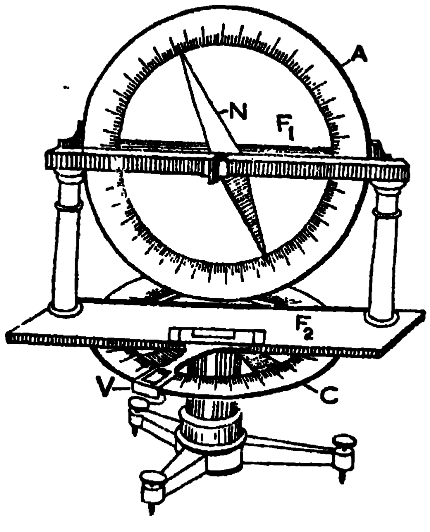

Fig. 272

The needle and the circle are mounted on a stand (F_1F_2) which can rotate about a vertical axis through the centre of the circle. The angle of rotation can be read off from a horizontal circular scale (C) with the help of the vernier V. The apparatus stands on three levelling screws and is made of non-magnetic material. A spirit level is generally attached to the turn-table F_2.

Experiment 48. *To measure the angle of dip at a place.*

Apparatus : Dip Circle.

Method: Remove all magnets and magnetic materials from the vicinity of the table where the work will be done. Level the instrument.

Turn the vertical circle about its axis until the needle points along the 90°—90° line. The needle is now vertical and the plane of the circle is perpendicular to the magnetic meridian*.

Now, turn the vertical circle through 90°. (This will bring its plane to the magnetic meridian. The needle can now rotate in the magnetic meridian and comes to rest in the direction of the earth's field.)

Read the positions of the ends of the needle on the vertical scale. Turn the vertical circle with the needle through 180° and take the same readings.

The mean value of these four readings gives the dip.

244. Both Poles of a Magnet have the same Strength

Float a magnet on a piece of cork in water (Fig. 273). The magnet turns and comes to rest in a fixed direction—roughly north and south. It does not move bodily towards the north or the south. If disturbed from the position of rest, it comes back to rest pointing in the same direction.

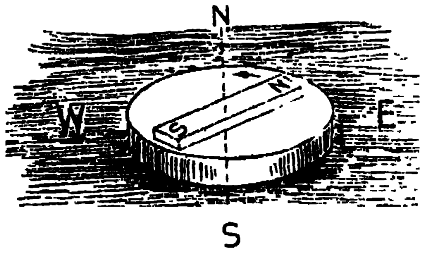

Fig. 273

This shows that the forces exerted on the poles of the magnet by the earth's field are equal and opposite. These two forces constitute a couple tending to rotate the magnet. If the force on one pole were greater than that on the other, the magnet would have bodily moved in the direction of the stronger force. But this does not occur.

Since the forces on the poles are equal, the poles must have the same strength.

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* The axis of rotation of the needle is in the magnetic meridian. The horizontal component of the earth's field is unable to produce a rotation of the needle. The needle can move only under the action of the vertical component and hence points vertically downwards.

245. Action of a Compass Needle

We are perhaps now in a better position to understand the action of a compass needle. The needle has a small cap fitted to it at its centre. It is pivoted at the cap on a nonmagnetic stand as shown in fig. 274. The needle turns in a horizontal plane under the action of the horizontal component of the earth's field. The vertical component tends to tilt the magnet in the vertical plane through it, but the effect is imperceptible.

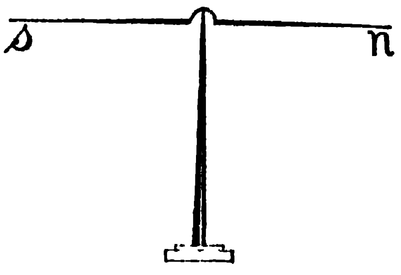


Fig. 274

246. Mariner's Compass

The directive porperty of a magnet led to its use for guidance in travel and navigation. The Chinese are reported to have used suspended lodestones for this purpose as far back as ten centuries B.C.

The modern magnetic compass is used by sailors and airmen for finding the true bearings of ships and aeroplanes. There are many models. The following is a simplified description of the essential parts.

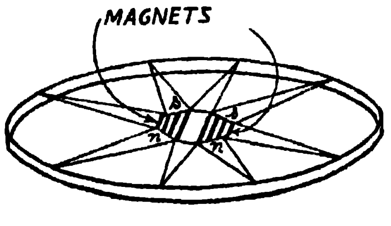


Fig. 275

A number of strongly magnetized, small, parallel steel needles, connected rigidly together, constitutes the system which, by its orientation, shows the magnetic north. This magnet system is attached to an aluminium ring by means of silk threads (Fig. 275). The ring is fixed to a disc with the 32 points of the compass marked on it (Fig. 276). The north and south points of the card lie exactly in the direction of the north and south poles of the magnet system.

In the centre of the

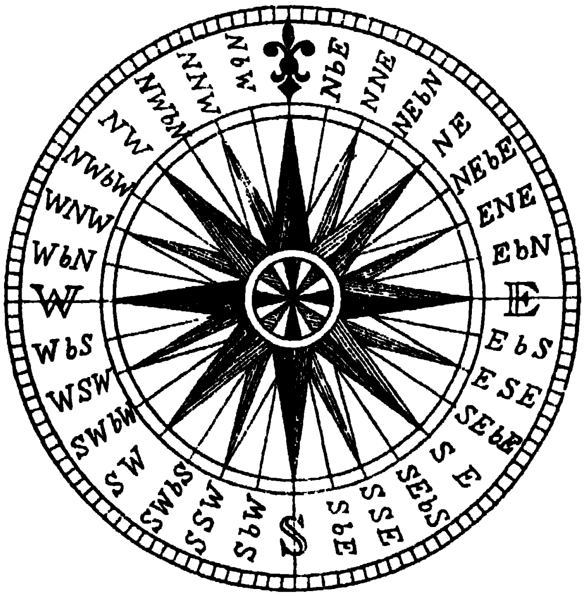


Fig. 276

disc there is an inverted sapphire cup (Fig. 277), which rests on a sharp iridium point P supported from the base of the compass bowl. The disc C, carrying the magnets (n, n shown in section) can turn on the pivot with a minimum of friction as both sapphire and iridium are very hard.

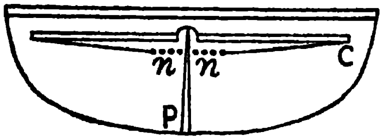


Fig. 277

A line laid on the glass cover of the compass bowl marks the fore-and-aft direction of the ship. As the ship turns the points on the disc below this line give the direction of motion of the ship relative to the magnetic north. If the declination at the place is known the direction of motion of the ship relative to the geographic north can then be easily found out.

The iron and steel in a ship produce a magnetic field, the effect of which is to make the compass needle deviate from the true magnetic north. This deviation is found by experiment before the ship starts on her journey and the effect is corrected by magnets placed near the compass.

247. Position of Poles of a Bar Magnet

Experiment 49. *To find the position of the poles of a bar magnet.*

Apparatus: Drawing board, pin and paper, two tall pins, thread, bar magnet, small magnetic needle (encased in glass), long compass needle.

Method. First, you have to determine the plane of the magnetic meridian.

For this purpose remove all magnets and magnetic materials from the neighbourhood of your working table. Take a long, thin compass needle and place it on the table. By means of two tall pins, placed fairly apart on the table, stretch a fine thread over the needle parallel to its axis (Fig. 278). The vertical plane through the thread is the magnetic meridian.

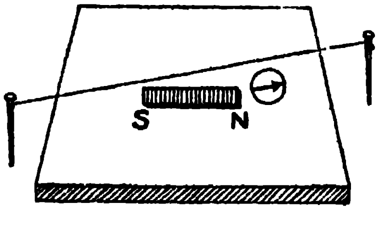


Fig. 278

Now, fix the drawing paper on the drawing board with pins and trace the outline of the bar magnet on the paper. Place the small compass needle close to one end of the magnet and turn the drawing board till the needle is parallel to the stretched thread. The needle will then be in the magnetic meridian.

(At this position of the needle the fields due to the earth and the two poles of the bar magnet are in the same direction. The effect of the far pole may be neglected due to its much greater distance compared with that of the nearer pole. Since the resultant field is in the direction of the earth's field, this is also th[illegible]ection of the field due to the nearer pole).

Mark the two ends of the small needle by two dots on the paper. Remove the needle to two other places near the same end of the magnet and repeat (Fig. 279). Join the pairs of points marking the ends of the needle at any position. When these lines are produced, they intersect at a point within the magnet. This is the position of the pole at the end considered.

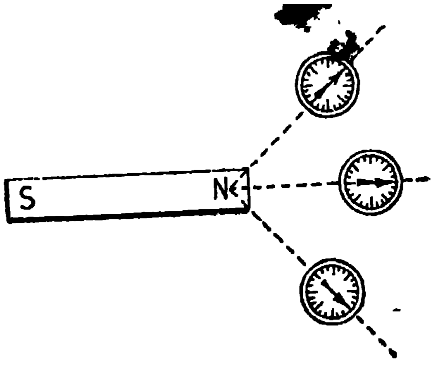
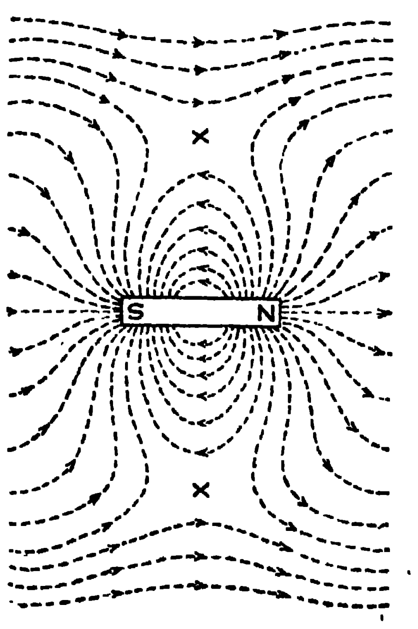


Fig. 279

Repeat the operations for the other end of the magnet.

248. Tracing the Lines of Force of a Magnet

In tracing lines of force of a magnet one has to remember that the magnet is surrounded by the earth's magnetic field. The field observed around a magnet is, therefore, the combined effect of the fields due to the magnet and the earth.



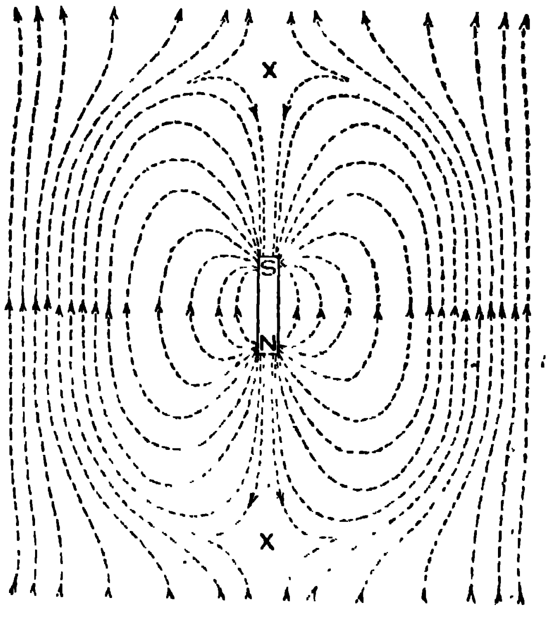


Fig. 280 North pole pointing north. Fig. 281 North pole pointing south.

This resultant field may be easily mapped by the method described in § 235. The pattern of the lines of force would depend on how the magnet is placed relative to the earth. Some patterns are given. The position of the magnet relative to the earth is stated under each.

249. Neutral Points

In each of the above diagrams you will find two points marked with a cross. There are no lines of force passing through the regions around these points. They are known as neutral points. The point at which the field due to one magnet is exactly equal and opposite to that due to another (here, the earth), is called a *neutral point.* There is no resultant field at a neutral point. A small compass needle placed at a neutral point will not orient itself in any favoured direction. It shows no directive property at a neutral point.

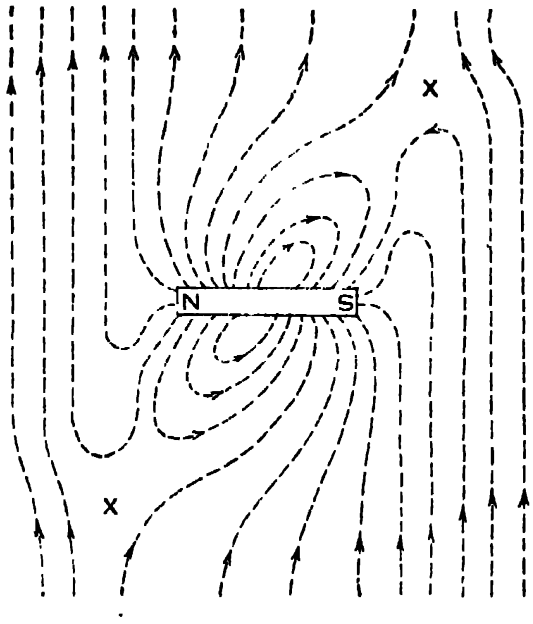


Fig. 282 North pole pointing west.

It may be interesting to compare the behaviour of an iron ring which has been magnetized by stroking round and round with one pole of a strong magnet. The ring is magnetized, but no free poles appear. When suspended it shows no directive effect. If cut, free poles appear (Fig. 283).

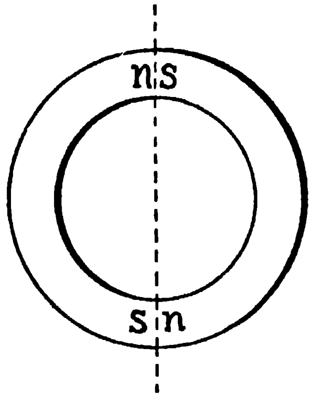


Fig. 283

Exercises

1. Why do we find it necessary to consider the earth as a huge magnet? What is the nature of the magnetic field due to the earth? Is the material of the earth's crust magnetized?

2. What do you understand by the *magnetic elements of the earth's field*? Define these quantities.

How would a freely suspended magnetic needle orient itself at the magnetic north pole?

3. How would you prove that both poles of a magnet have the same strength?

4. Draw a neat sketch of a mariner's compass and add a short description.

How is it used to find direction?

5. What is a neutral point? How does a magnetic compass needle behave when placed at such a point?

CHAPTER 30

ELEMENTARY FACTS ABOUT STATIC ELECTRICITY

250. Introductory

Within the past century man has discovered hundreds of ways of putting electricity to his use. In our homes we see lamps, fans, heaters, the radio, the telephone, all operated by electricity. In industry electricity plays an even more important role. It operates all kinds of machines in factories, and does thousand other jobs for us.

Where does electricity come from? You must have seen the high transmission lines which carry electricity to our homes, offices and factories. If you follow those lines back to the source you will come to large buildings, called *power stations*, which house machines, named *generators*. These machines are operated by thermal or water power, and convert a part of the energy they receive into electricity.

In the following pages we shall examine some of the properties of electricity.

251. Electricity from Friction

Experiment 50. (*i*) Rub a clean, warm glass rod with silk and hold it near small pieces of paper, straw, hair etc. They will be attracted by the rod (Fig. 284). The rod is said to be *electrified* or have acquired an *electric charge.*

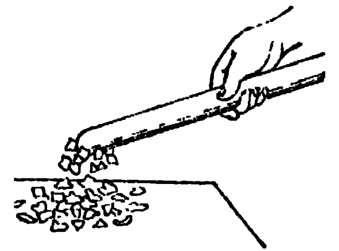


Fig. 284

(*ii*) Repeat the experiment, using an ebonite or sealing-wax rod rubbed with flannel or catskin, or your fountain pen rubbed with wool.

(*iii*) Balance a metre scale on an inverted watch glass and bring a charged ebonite rod near one end of it.

Charges of electricity are produced by friction in this way. They remain on the bodies rubbed and attract nearby light bodies. The ancient Greeks (about 600 B.C.) knew that amber rubbed with cloth could attract dust particles, small pieces of paper, feather, straw etc.

While combing hair with a hard-rubber comb on a dry winter day, you might have noticed that the comb attracted hair. If your hair is dry enough you may hear the crackle of electric sparks as you run the comb. In a dark room you may even see these sparks.

Careful experiments show that when two substances are rubbed together, or even brought into close contact, each acquires an electric charge. It is not, however, always easy to demonstrate the effect, as with most substances the effect is rather small.

Under favourable conditions charges due to friction may accumulate on bodies and prove potentially dangerous, so much so that precaution has to be taken against such accumulation. On trucks which carry petrol, friction caused by motion builds up electric charges. If these charges were allowed to accumulate, they would cause a spark which might ignite the petrol vapour and start a fire. To prevent the accumulation of charge a small chain is kept hanging from the frame of the truck and drags on the ground. The charge passes into the earth through this chain.

Frictional electric charges build up on the moving belts that carry power to machinery in factories The machines to which the belts are connected must be well earthed to prevent accumulation of charge on them.

Clouds carry large static charges, which give rise to lightning (See § 262).

252. Two Kinds of Electricity

Experiment 51. (*i*) Rub a hard-rubber or ebonite rod with flannel to charge it. Suspend it from a fixed support by a silk thread (Fig. 285). Bring near one end of it another charged hard-rubber or ebonite rod. They repel each other.

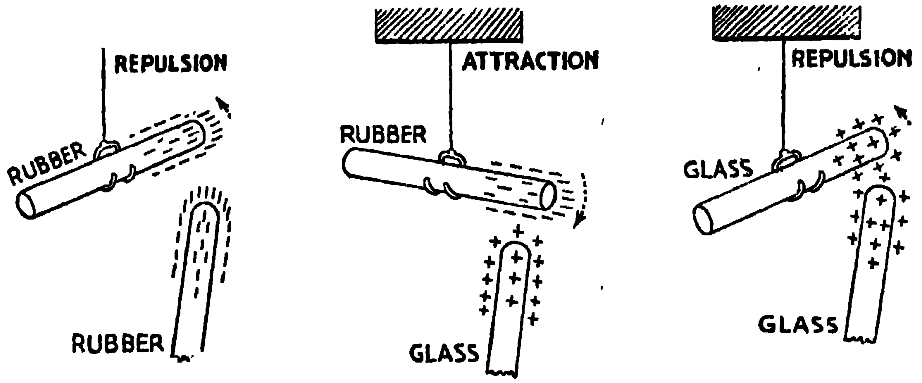


Fig. 285

(*ii*) Repeat the experiment using two charged glass rods. They also repel each other.

(*iii*) Suspend the charged hard-rubber rod as before. Then bring near it a charged glass rod. They attract each other.

From these experiments we conclude that

(*a*) *There are two kinds of electricity.*

(*b*) *Like charges repel, while unlike charges attract, one another.*

The kind of electricity as on glass has been called *positive*, and the other (as on ebonite), *negative*. The justification of the names lies in the fact that when equal amounts of the two kinds are put together, no electrical property is exhibited; one neutralizes the effect of the other.

253. Electrons and Protons

To understand why there is attraction in one case and repulsion in another it will be useful for us to remember how matter is constituted. Matter is made up of atoms. Each atom has a nucleus around which move *electrons* at various distances and in various paths, much like the planets around the sun. The nucleus has two kinds of particles, the *proton* and the *neutron*.

Electrons are the tiniest particles known to us. Each electron has the same amount of negative charge and a mass of 9×10^{-28} gm. A smaller charge than that on the electron is not known to exist. This charge, called the *electronic charge*, is the unit of negative charge.

Protons have the same amount of electric charge as electrons, but the charge is positive in nature. This is the smallest known charge of positive electricity. A proton is over 1800 times heavier than an electron.

Neutrons have no electric charge. A neutron is very slightly heavier than a proton.

Explanation of electrification by friction. Each atom normally has the same number of electrons and protons. It is, therefore, *electrically neutral*. This picture of the atom leads to a simple explanation of electrification by friction. When the glass rod was rubbed with silk electrons were pulled off the atoms on the surface of the glass. This left the glass with an excess of positive charge as it then had more protons than electrons.

When the ebonite rod was rubbed with flannel electrons were removed from the flannel by the rod. This left the flannel with an excess of protons, i.e., with a positive charge, and the rod with an excess of electrons, i.e., with a negative charge. It is clear that the charges on the ebonite rod and flannel must be equal and opposite. This also holds for glass and silk, in fact, in all cases where electricity is produced by friction. This is also confirmed by experiment.

254. Conductors and Insulators

A conductor is a substance that allows electricity to move along it. Metals are good conductors. Good conductors of heat are generally good conductors of electricity also. The fact that a metal is a good conductor of electricity explains why it is difficult to demonstrate production of electric charge on it by friction. The charges produced flow along the metal into the person holding it, and thence into the earth.

An insulator (or a non-conductor) is a substance which does not allow electricity to move along it. Rubber, glass, porcelain, sulphur, mica, ebonite, dry air etc. are examples of good insulators. To prevent the leakage of electricity from telegraph, telephone, transmission or supply lines, the wires are supported on glass or porcelain knobs.

There is no sharp line separating conductors from insulators. Most substances conduct a little; there is no perfect conductor nor perfect insulator. Dry air is the best insulator and silver the best conductor, copper being a close second. Moist air carries away the charge from a body quickly. This explains why it is difficult to carry out experiments on static electricity in our country except during the winter.

255. Electroscopes

An electroscope is an instrument for detecting the presence of an electrical charge on a body and for determining its nature, positive or negative.

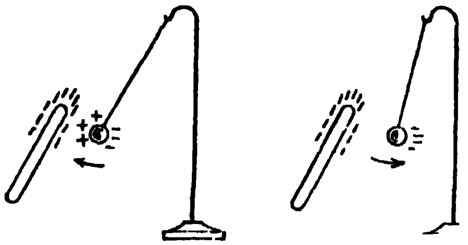


Fig. 286

Pith-ball electroscope. One of the earliest forms of electroscope consists of a small ball of dried pith suspended by a silk thread from a support (Fig. 286). Its action is improved by coating the ball with a thin foil of metal, say, aluminium.

Bring a charged rod near a pith-ball electroscope. The ball will be attracted towards the rod. If the ball touches the rod, it will be repelled, since, by contact, it acquires the same kind of charge as on the rod.

The ball may be charged with a known kind of electricity in this way. The nature of the charge on another body can then be tested by bringing it near the ball *gradually from a distance.*

If there is attraction, the charges are of opposite kind. If there is repulsion, the charges are similar.

Gold-leaf Electroscope. Much more sensitive than the pith ball is the gold-leaf electroscope. It consists of two narrow strips of very thin gold leaf (now, more commonly, aluminium leaf) *L*, *L* (Figs. 287 and 288) attached to a metallic rod *R* which projects through the insulating stopper *S* of the containing vessel *V*. *V* prevents disturbance of the leaves by air currents. It is made of metal (Fig. 287) or glass coated with metal foil (*M*, Fig. 288), and has glass front and back. The rod *R* usually terminates in a metal knob or in a flat plate (*C*). A small quanity of a drying agent, such as phosphorus pentoxide, may be kept within the vessel to maintain a dry atmosphere in it. The metal part of *V* may be connected to the earth as in fig. 287. In place of gold leaves, an aluminium leaf hung beside a brass strip is often employed. The stopper *S* is made of sulphur, paraffin or some other good insulating material.

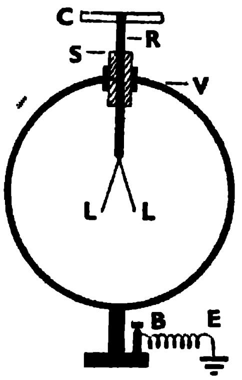


Fig. 287

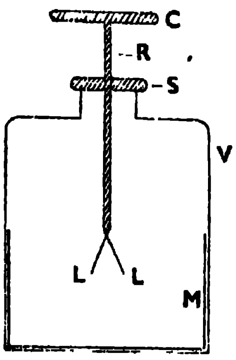


Fig. 28

Charging an electroscope by conduction

To charge an electroscope by conduction, a body charged with a known kind of electricity is placed in contact with the plate *C* of the electroscope and then removed. The charge transferred to *C* spreads over *C*, *R* and *L*. Due to the repulsion of like charges on *L*, the leaves diverge.

To test the nature of the charge on a body

When a charged body is gradually brought nearer to a charged electroscope, the divergence of the leaves will increase if the charges are of the same kind; but diminish if they are of opposite kinds. When the charged body is removed without having at any time touched the electroscope, the divergence of its leaves is restored to the initial value.

A charged electroscope slowly loses its charge by leakage, and the leaves finally collapse.

256. Electrostatic Induction

The two kinds of electricity in a body undergo a relative displacement under the influence of an electric charge. The phenomenon is known as *electrostatic induction*. It is very pronounced in metals and may be demonstrated as follows.

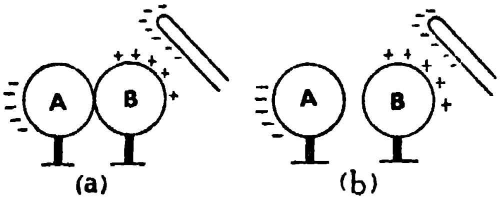


Fig. 289 (a) & (b)

A and *B* are two insulated metal balls mounted on separate stands, but are initially in contact (Fig. 289a). A charged rod, say, of ebonite rubbed with fur, is brought near *B*. With the rod held in position, *A* is separated from *B* (Fig. 289b). The rod is *then* removed.

On testing *A* and *B* separately, each will be found to have acquired an electric charge but the charges on them are of opposite kinds. *A*, the remote conductor acquires the same kind of charge as on the rod. *B*, the one near the rod, acquires a charge of the opposite kind.

The charge on the rod is called the *inducing charge*, and the rod itself, the *inducing body*. Charges on *A* and *B* are called *induced charges*.

The explanation is simple. The ebonite rod has a charge of negative electricity. It repels the electrons on the conductor placed near it. Since electrons are free to move in a conductor, they move to the remotest end on the latter and produce an excess of negative charge at that place. The part of the conductor near the rod has a deficit of electrons and shows a positive charge.

In general, the induced charge having the same sign as the inducing charge moves farthest away from the latter. Because of this tendency to move as far away as possible it is called *free charge*. If the body were connected with the earth the free charge would escape into the earth.

The other induced charge of the kind opposite to the inducing charge, is held nearest to the latter by attraction and is known as *bound charge*.

257. Charging a Gold-leaf Electroscope by Induction

To charge a gold-leaf electroscope by induction we may proceed as follows:

(*i*) Take an ebonite rod rubbed with fur if the electroscope is to be charged with positive electricity. To charge with negative electricity take a glass rod rubbed with silk. (Note that the inducing body must have a charge opposite in kind to that with which you want to charge the electroscope.)

(*ii*) Bring the inducing rod close to the cap of the electroscope (but the two must not touch).

(*iii*) Momentarily connect the cap with the earth while the inducing rod is held in position.

(*iv*) Remove the connexion between the earth and the cap (the inducing rod continuing in position).

(*v*) Remove the inducing rod.

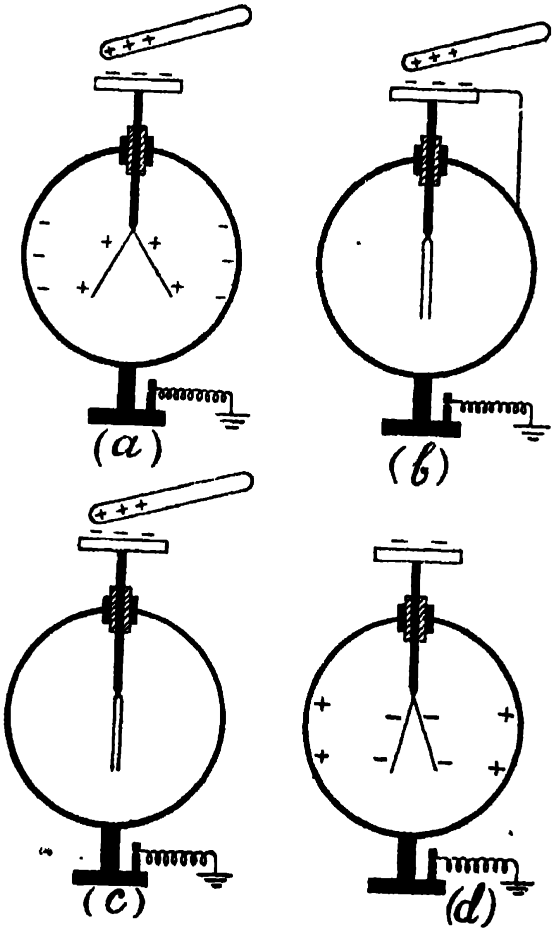


Fig. 290 (a) to (d)

Steps (*ii*) to (*v*) are illustrated in Fig. 290 (*a*) to (*d*) with a positively charged inducing rod, charging the electroscope with negative electricity. The metal part of the casing may profitably be kept earthed throughout as shown in the figures.

Fig. 290(*a*) shows that while the cap of the electroscope acquires negative electricity by induction the leaves acquire a positive electricity. (The charge on the leaves induces an opposite charge on the inner side of the metal casing).

Fig. 290(*b*) shows the effect of connecting the cap to the earth, here through the metal casing. (Touching the cap with a finger also connects it to the earth through the body of the demonstrator unless the latter is insulated.) The free charge escapes to the earth, and the leaves collapse.

Fig. 290(*c*) shows that though the earth connexion has been removed, the leaves continue in a collapsed state, there being no charge on them.

Fig. 290(*d*) shows the final state when the inducing rod has been removed. The bound charge, so long held on the cap by the attraction of the charge on the inducing rod, now spreads over the cap, the connecting rod and the leaves. As a result (i.e., due to the repulsion of the charge on the leaves) the leaves diverge.

The charge on the leaves induces a charge on the casing. The bound charge will be on the inner side of the casing. Its attraction for the opposite charge on the leaves will tend to maintain the divergence of the leaves. Remember that the cap, connecting rod and leaves (*C*, *R*, *L*; Figs. 287 and 288) are insulated from the metal casing i.e., the metal part of the containing vessel.

258. More Facts about Induction

Equality of charges. When electrification is produced by induction or friction, equal amounts of opposite charges appear. This is easy to understand as electrification arises out of a transfer of electrons from one place to another. The region from which electrons move away is left with an excess of positive charge of the same amount.

Earlier you have read about the principle of conservation of energy; energy cannot be created or destroyed. In chemistry, you must have read of the principle of conservation of matter; matter cannot be created or destroyed. Here you may add a third principle—the principle of conservation of electricity. It states that electricity cannot be created or destroyed. In all electrical phenomena, appearance of electricity of one kind at any place is always accompanied by the appearance of an equal amount of the opposite kind of electricity at some other place. The production of a charge necessitates the production of an equal opposite charge. The charges are separated—not created.

Relation between inducing and induced charge. When induction takes place the charge induced on a body is generally smaller in amount than the inducing charge. When, however, a conductor fully surrounds the inducing charge, the charge induced on the conductor will be equal in amount to the inducing charge. This was established by Michael Faraday, one of the pioneers in electricity.

Induction precedes attraction. An uncharged body is attracted by a charged body. You have already seen how a charged glass rod attracts uncharged pieces of paper etc.

When the charged body approaches the uncharged one, induction takes place, an opposite charge being induced on the side of the uncharged body nearest to the charged one. An attraction ensues between these unlike charges.

The like charge induced on the body is farthest from the inducing charge. Hence the repulsion between these like charges is less than the attraction. The net result is therefore an attraction between the two bodies. *The resultant attraction is an effect* *of induction.* This is what we mean by saying induction precedes attraction. The similarity with magnetism may be noticed.

Repulsion is a surer test of electrification. When attraction takes place between two bodies one of which is known to be electrified, the other may have an unlike charge on it or no charge at all. But if there is repulsion between one charged body and another, the latter must have a like charge. Repulsion is, therefore, a surer test of electrification. Here also the behaviour is the same as in magnetism.

259. Proof Plane

A proof plane is a small metal disc with an insulated handle. When the disc is made to touch a charged body it is itself charged by conduction with the same kind of electricity as on the body. If the disc of the proof plane be held near a charged electroscope by the insulated handle, the leaves of the electroscope will diverge more if it has the same kind of charge as on the proof plane. If the charges on the two are of opposite kinds, the divergence will diminish.

260. Distribution of Charge on a Conductor

When an electric charge is given to an insulator, the charge resides where it is placed. As, however, no insulator is perfect the charge slowly spreads to neighbouring areas and penetrates to some extent into the interior.

A charge placed on a conductor behaves differently as charges can move very easily along conductors. Due to mutual repulsion particles of electricity on a conductor try to move away from each other as much as possible. This would lead us to expect that on a conductor the charge will normally reside on the outer surface.

Electricity resides on the outer surface of a conductor. The following experiment, due to Faraday, may be performed in support of the statement. Take a conical linen bag which is fastened at its wider end to an insulated metal ring. It has a silk thread attached to its apex to draw it inside out (Fig. 291). Linen is a fairly good conductor. Charge the bag with electricity. With the help of a proof plane and electroscope it is easy to show that there is a

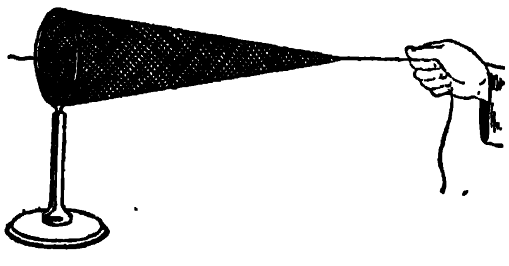


Fig. 291

charge only on the outside of the bag, and none on the inside. If now the bag is turned inside out, it will be found that the charge has moved to the new outer surface which was erstwhile the inner surface.

Faraday's chamber experiment. To show that charges reside wholly on the surface of conducting bodies, Faraday performed a classic experiment. He built a cubical box (of sides 12 ft. each), covered it with tinfoil and shut himself up in the box taking with him various instruments for detecting electricity. The box was then charged highly with electricity. Faraday did not find any trace of electricity in the interior although large sparks were darting forth from every part of the outer surface of the cage.

The effect can be demonstrated in the laboratory by placing an electroscope inside a cage of wire gauge kept on an insulating support (Fig. 292). On bringing a charged rod close to the cage, or even charging the cage electrically from an electric machine, it will be found that the electroscope is unaffected.

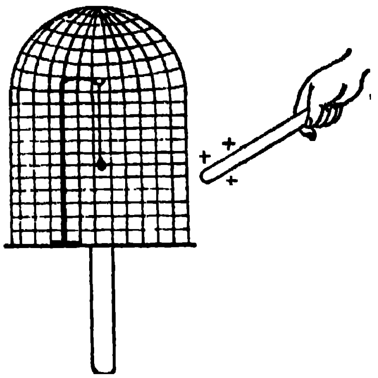


Fig. 292

Electrical screening. These experiments show that a region can be screened or shielded from external electrostatic influences by a continuous metallic cover or even a metallic net-work. The effect so produced is known as *electrical screening* or *electrical shielding.*

261. Effect of Sharp Points on the Distribution of Electricity on Conductors

When a charged conductor has a sharp point on its surface, it is found that there is a high concentration of electricity at the point. This concentration produces interesting and important effects. Molecules of air in the vicinity of the point are acted on inductively by the charge on the point and attracted towards it. On contact the induced opposite charge is neutralized. This produces *two* effects, viz:

(*i*) The charge on the point is reduced. The conductor gradually loses charge in this manner. This action is known as the *discharging action of points.*

(*ii*) The attracted air molecules, on contact with the point, acquire the same kind of charge as the point has, and are then repelled. The repelled molecules may be sufficient in number and speed to produce a wind, called an *electric wind*, strong enough to deflect the flame of a candle (Fig. 293).

When concentration of charge at such points is very high, electrons may be torn off by the charge from air molecules. When this takes place a glow may be observed near the point. This is known as *brush discharge*.

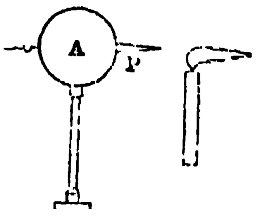


Fig. 293

262. What is Lightning

Falling water-drops, formed by the condensation of moisture in a cloud, break up under certain conditions into unequal fragments. In the process of splitting the larger fragments generally acquire a positive charge, and the smaller fragments, a negative one. During a thunderstorm the warm air near the surface of the earth rises, expands and cools. Its uprush not only splits the water-drops, but also brings about a separation of charge, the smaller drops with the negative charge being carried upward. A charged cloud thus comes to be formed, with negative electricity at the top and positive electricity at the bottom. Portions of the cloud with only one kind of charge may get separated.

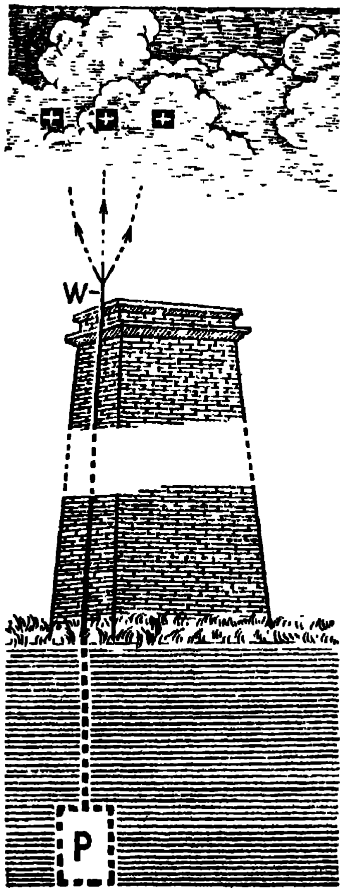


Fig. 294

When large charges have been built up in this way, the attraction between opposite charges may cause an electric discharge to pass between two parts of the same cloud, or between two clouds, or between a cloud and the earth. A discharge of this kind is called *lightning*. The path followed by the discharge is often marked by a brilliant luminosity. The loud noise accompanying a flash of lightning is called *thunder*. It is due to the sudden expansion of the air along the path of the lightning.

263. Protection from Lightning

A charged cloud will induce an opposite charge on the ground. The attraction between the opposite charges may cause lightning to strike the earth at a point which is *electrically nearest* to the cloud. The tallest structure in a locality is, therefore, most likely to be struck by lightning.

To protect a tall building from lightning a metal rod, called a *lightning rod*, with several points is placed on the roof. A strong metal conductor passing over the outside of the house connects the rod to the moist earth (as shown by the plate *P* in fig. 294). The cloud discharges into the earth through this conductor. The points on the conductor help this discharge.

A conductor of height *h* protects a conical region of radius 4*h*.

In a lightning storm, safe places of shelter are (*i*) a steel-frame building, (*ii*) a building well provided with lightning conductors, (*iii*) an automobile or a shed with an earthed metal roof, (*iv*) a cave, dugout or ravine. One should keep away from wire fences, isolated trees, walls, stoves and fire places.

Exercises

1. What are electrons and protons? Explain electrification by friction in terms of electrons and protons.

2. Why do we say that the production of a charge necessitates the production of an equal opposite charge?

Electricity is said to obey a law of conservation. State any other similar law that you may know.

3. Draw a neat sketch of a gold-leaf electroscope and label the parts. How will you charge it with positive electricity (*a*) by conduction, (*b*) by induction ?

4. You are given a gold-leaf electroscope, a glass rod and a piece of silk, and asked to determine the kind of electricity which will be produced on a stick of sealing wax when rubbed with flannel. State clearly what you will do.

5. Explain the following statements:—

(*a*) Repulsion is a surer test of electrification than attraction.

(*b*) Induction precedes attraction.

(*c*) The charge on a conductor resides on the surface.

6. Explain the action of a point in discharging a conductor. How does a lightning rod act?

7. What is lightning? Why is it not safe to stand in the open during a thunderstorm? What precaution will you take in such a situation?

ELECTRIC FIELD AND POTENTIAL

264. Coulomb's Law

Coulomb was the first to establish quantitatively the law of force between electric charges. By 1777 he had devised an apparatus known as the *torsion balance.* With its help he established the result that *the force between two (point) charges varies directly as the product of the charges and inversely as the square of the distance between them.* This is known as **Coulomb's law of inverse square** in electrostatics.

Mathematically, the law is expressed by writing

$$f = \frac{q_1 q_2}{kr^2} \qquad (264\cdot1)$$

where f is the force between two point charges of magnitude q_1 and q_2 separated by a distance r. k is a constant, called *dielectric constant,* whose value depends on the nature of the medium separating the charges. k for vacuum is taken as unity. For air it has practically the same value.

The electrostatic unit of charge. Formula 264·1 helps us in defining the electrostatic unit of charge. *An electric charge which repels an identical charge at a distance of one centimetre from it in vacuum with a force of one dyne is said to have a magnitude of one electrostatic unit.* There is no commonly accepted name for this unit. We shall refer to it as the e.s.u. (abbreviation for electrostatic unit) of charge. The practical unit of charge is called the *coulomb.* It is equal to 3×10^9 e.s.u.

The charge on an electron or a proton is $4{\cdot}8 \times 10^{-10}$ e.s.u. It requires $2{\cdot}08 \times 10^9$ electrons to make up one e.s.u. of negative charge and as many protons for one e.s.u. of positive charge.

265. Electric Field

The region surrounding an electric charge where its influence may be detected is called an *electric field.* To investigate an electric field we may imagine a unit, positive, point charge to be placed in the field and find the force it would experience at different points of the field. This charge is called a *test charge.* Its presence in a field is supposed not to alter the field in any way.

The **intensity**, (or **strength**) **of an electric field** at a point is defined as the force in dynes which would act upon a unit positive test charge if placed at the point. The direction of this force is called the direction of the electric field. Intensity of the field is clearly a vector quantity. It is measured in dynes per unit charge.

If a charge of q *electrostatic units* be placed in a field at a point where the intensity is E *dynes per unit charge,* the force f on the charge q is given by

$$f \text{ (in dynes)} = E \times q$$

A point charge q produces at a distance r from it in air an electric field of intensity given by

$$E = \frac{q}{r^2}.$$

266. Lines of Force in an Electric Field

If a *small, free, positive* charge were released in an electric field it would be urged in a definite direction at every point of the field. *A line of force in an electric field is the path which a free, positive, point charge would describe if it moved everywhere in the direction of the field.* The tangent to a line of force gives everywhere the direction of the field.

Lines of electric force are assumed to originate on positive charges and terminate on negative charges. Attraction and repulsion between charges may be visualized if we endow the lines of force with the properties of a stretched, elastic cord. In fact, magnetic and electric lines of force have exactly similar properties, and all that we said about the former in § 234 will apply to the latter if we replace the magnetic terms by the corresponding electric terms.

267. Electric Potential

The concept of potential plays a very important role in physics, particularly in electrical phenomena. It may be introduced in the following way.

Consider a region of space initially free from all electrical influence and devoid of matter. Imagine a small charge $+q'$ placed anywhere in the region. It may be moved about in the region without any work of electrical origin being done, since there is no electrical force acting on it.

But let a charge $+q$ be placed at any point P in this region before the charge $+q'$ is brought in. To move $+q'$ from one point to another in the region *will now require work,* for the charge $+q'$ will be acted upon by the repulsive force due to $+q$. To move $+q'$ nearer to $+q$ some external agent will have to do work against the repulsion, while if $+q'$ moves away from $+q$ the repulsive force does work. Work is done also when either or both of q and q' are negative.

We thus find that the presence of a charge q confers upon the region around it a property in virtue of which work must be done in moving another charge from one point to another in the region, or in bringing this other charge from outside the region to a point within the region. This property is called *electric potential.* Its value at any point is measured by the work done (by an external agent) in bringing a unit positive charge from infinity (i.e. outside the region) up to the point.

If W units of work are required to bring q units of charge from infinity to a point P, the potential V at P is given by $V = W/q$ or $W = Vq$, i.e.,

Work = Charge × potential.

Potential difference. If the potential at a point A is V_a and that at a point B is V_b, then $V_a - V_b$ is the difference of potential between the points. The work W done in transferring a charge q from B to A will be given by

$$W = (V_a - V_b) \times q.$$

The Volt. The volt is the practical unit of potential difference. Two points are said to be at a potential difference of one volt when one joule of work is done in transferring one coulomb of charge from one point to the other.

Positive electricity tends to flow from places of higher to places of lower potential. It will be profitable for a beginner to remember that positive electricity tends to move from places of higher potential to places of lower potential. Negative electricity does the reverse.

When two bodies at different potentials are brought together, positive electricity will flow from the body at the higher potential to the body at the lower potential even if the latter has more electric charge on it. The flow continues until the potentials are equalized.

Analogy between potential and 'level' in hydrostatics. Potential may be compared with 'level' in hydrostatics. In fig. 295 the level of water in A is higher than that in B. When the stopcock D is opened water flows from A to B even though B may contain more water. The flow continues so long as the level in A remains higher than that in B, and stops when the *levels are equal.* The behaviour of the liquid is the same as that of a positive electric charge. When

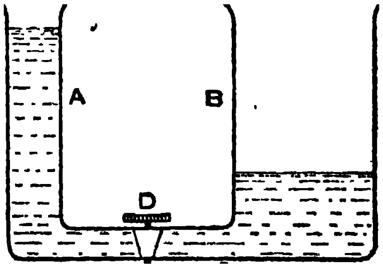


Fig. 295

two bodies at different potentials are connected together positive charge flows from the body at the higher potential to that at the lower until their potentials are equalized.

Hence potential behaves in the manner of hydrostatic pressure and may be called an 'electrical pressure' forcing a positive charge to move from higher to lower potential. Negative charge, however, behaves oppositely.

Analogy between potential and temperature. When two bodies at different temperatures are brought together, heat flows from the body at the higher temperature to that at the lower even if the latter has more heat in it. The flow continues until the temperatures are equalized. The behaviour of positive charge is like heat, and potential behaves in the manner of temperature.

268. Capacity

When a charge (q) is placed on an insulated conductor, it acquires a potential (v). The ratio of the charge to the potential is called the capacity (C) of the conductor. In symbols,

$$C = \frac{q}{v}.$$

Capacity may, therefore, be defined as the charge required to raise the conductor to unit potential.

The capacity of a conductor depends on (*i*) its area, (*ii*) the medium surrounding it and (*iii*) the presence of neighbouring conductors.

In general, capacity increases with area. In the case of spheres, capacity is proportional to the radius.

Capacity increases if the conductor is surrounded not by air, but by some nonconducting solid or liquid medium such as paraffin wax, shellac, sulphur, glass, oil etc.

The presence of a neighbouring conductor, particularly if it is earthed, increases the capacity of a conductor. (See below.)

269. Condensers

Two neighbouring conductors, one insulated and the other ordinarily earthed, with a non-conducting medium separating them, constitute an electric condenser. It is used for storing electric charge. Practical condensers are generally constructed to have simple geometrical shapes. Thus, in parallel-plate condensers the conductors are two parallel plates, in spherical condensers they are two concentric spherical surfaces and in cylindrical condensers, the conductors are two coaxial cylinders.

Principle of the condenser. A condenser may be looked upon as a device in which the capacity of a conductor (the insulated one) has been increased by the presence of a neighbouring earthed conductor. How this increase is brought about may be understood from the following diagram (Fig. 296).

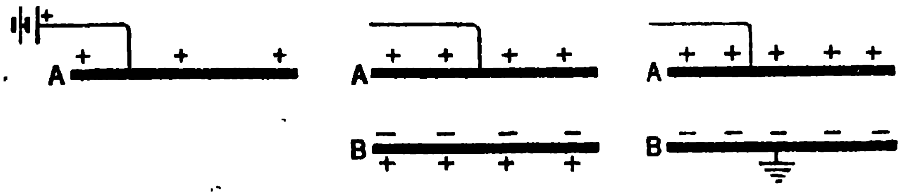


Fig. 296

Suppose the metal plate A is joined to the positive pole of a battery. It will acquire a positive charge and be at the potential of the pole. When another plate B is brought near A, it is charged by induction as shown. The negative charge on B brings more positive charge on A by attraction. This action is, however, opposed by the positive charge on B. When B is earth-connected the free positive charge on it escapes into the earth, leaving the negative charge on B free to attract yet more charge on A. A is still at the potential of the pole, but holds more charge. Its capacity has increased.

The two plates of a condenser have equal and opposite charges. The capacity of a condenser is defined as the ratio of the charge on either conductor to the difference of potential between the conductors.

If C = capacity, Q = charge and V = potential difference,

$$C = Q/V \quad \text{or} \quad Q = CV$$

i.e., charge = capacity × potential difference.

The practical unit of capacity is the farad. A condenser in which a charge of one coulomb on either conductor causes a potential difference of one volt between the conductors is said to have a capacity of one **farad**.

$$1 \text{ farad} = \frac{1 \text{ coulomb}}{1 \text{ volt}}$$

A farad is too large a capacity for most practical purposes. It is therefore common to use the microfarad, which is one millionth of a farad.

1 microfarad ($1\ \mu f$) = 10^{-6} farad.

270. Factors affecting the Capacity of a Condenser

The capacity of a condenser depends on (*i*) the size of the conductors, (*ii*) the distance between them and (*iii*) the nature of the medium separating them.

Capacity increases with the size of the conductors. It also increases as the distance between the two conductors *diminishes*.

Other factors remaining constant it has been found that the capacity is materially affected by the medium filling the space between the two conductors.

The influence of a non-conducting medium on the capacity of a condenser is expressed through a constant called *dielectric constant* or *specific inductive capacity*. It is the ratio of the capacity of a condenser when the given medium separates its conductors to the capacity when air (more strictly, vacuum) separates them. This constant is usually denoted by the symbol k (latin kappa). Values for some media are given below.

Substance	k	*Substance*	k
Air (at N.T.P.)	1·000586	Shellac	3—3·7
Paraffin wax	2·0—2·3	Sulphur	3·6-4·3
Dry paper	2·0—2·4	Paraffin oil	4·6-4·08
India rubber	2·1—203	Mica	5·7-7
Wood	2·5—2·7	Flint glass	7—10

271. The Leyden Jar

One of the earliest forms of condenser is the Leyden jar, named after its place of invention (1745) in Holland. It is a glass jar (V; Fig. 297), the bottom and about three-quarters of the sides of which are coated inside and outside with tinfoil. The inner and the outer tin-foils (M_1, M_2) are the two 'plates' of the condenser with glass as the medium separating them. A brass knob K is attached to one end of a brass rod R passing through an insulating lid S fitting the mouth of the jar. Electrical connection between the knob and the inner coating of the tin-foil is made by a short length of brass chain hanging from the rod. The jar is coated on the outside with shellac to reduce leakage of electricity.

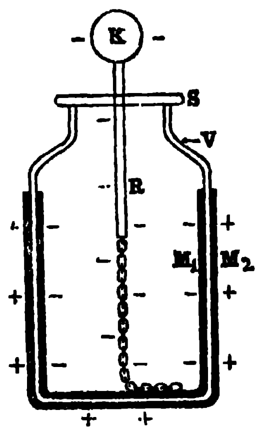


Fig. 297

To charge the jar it is held in the hand by the outer coating, which connects it to the earth. The knob is then brought into contact with one terminal of an electrostatic generator, say, a Wimshurst machine. Fig. 297 shows the distribution of charge when the knob is connected with the negative pole of the machine.

The jar is discharged by a pair of tongs, as shown in fig. 298. It is of brass with an insulating handle. One knob of the tong is laid against the outer coating and the other gradually brought near the knob *K* of the jar. The discharge occurs with a bright spark.

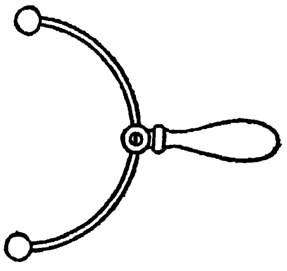


Fig. 298

The capacity of a Leyden jar depends on the area coated on both sides with metal foil and the thickness of the glass between foils. Capacity is directly proportional to the area and inversely proportional to the thickness.

A Leyden jar charged by an electrostatic machine may produce a dangerous shock when the two foils are simultaneously touched. There is a record of the discharge from a big jar passing through a rank of 180 soldiers making them all jump at once.

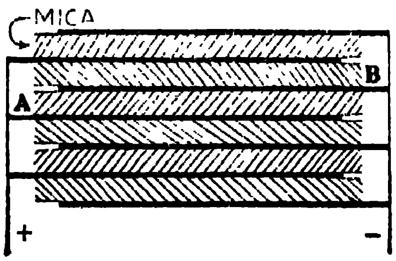


Fig. 299

Condensers with only one pair of plates will have a small capacity. Compact condensers of larger capacity are constructed of alternate sheets of tin-foil (thick lines in fig. 299) and paraffined paper or mica. Alternate tinfoils are connected together and form one terminal of the condenser (*A* and *B* of fig. 299). Parallel plate condensers of which the capacity can be varied by turning a knob are used for tuning radio receivers.

Exercises

1. State Coulomb's law of force between two point charges. Define the electrostatic unit of charge. What is the practical unit? How are the two units related?

2. What is an electric field? Explain the term 'test charge' in connection with an electric field.

Define 'intensity of a field'. What do you understand by direction of a field?

3. What are lines of electric force? What are the properties ascribed to them?

4. What do you understand by electric potential? What is its measure? Define the volt.

5. State the points of similarity between electric potential and temperature.

6. In what way can electric potential be treated as a pressure? State any similarity it may have with hydrostatic level.

7. What is a condenser? How does it act? Illustrate your answer with the Leyden jar.

8. What is meant by the capacity of a condenser? What is the practical unit of capacity and how is it defined?

Mention the factors on which the capacity of a condenser depends, and state the nature of the dependence.

What do you understand by the 'dielectric constant' or 'specific inductive capacity' of a medium?

CHAPTER 32

NATURE AND PROPERTIES OF ELECTRIC CURRENTS

272. The Electric Circuit

The benefits we derive from electricity follow not from electric charges at rest, but from electric charges in motion, i.e., from the electric current. It will be easy for us to understand some of the basic facts about electric currents when we remember the behaviour of a current of water.

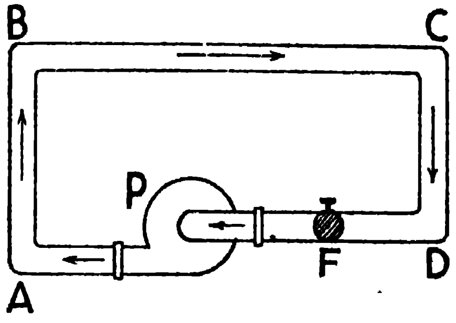


Fig. 300

Let us see what we will have to do to maintain a current of water in a system of pipes *ABCD* as in fig. 300. The pipes must be arranged to form a *closed path* so that water may flow continuously as long as we desire. Besides, there must be a pump *P* somewhere in the closed path to force water through the pipes and maintain the circulation. To turn the current on and off at will we may also place a stop-cock *F* somewhere in the path.

This simple arrangement tells us what we will have to do *to maintain* an electric current. It is as follows :

(*i*) There must be a closed path along which electricity will move. Since insulators do not allow motion of electricity along them, the closed path must be made up of conductors. The conductors are generally metallic wires. Such a path is called an electric circuit.

(*ii*) Somewhere in the circuit there must be a device corresponding to the pump which will force electricity through the circuit and maintain the circulation. The cell of a torch, or a dynamo is such a device. It is called a source of electromotive force, (abbreviated e.m.f.). E.m.f. has been discussed in §. 276 later.

(*iii*) To turn the current on and off at will we should include a suitable device in the circuit. It is called a *switch*. The position of the switch or the position of the source of e.m.f. in the circuit is immaterial, just as the positions of *P* and *F* in fig. 300 are.

273. What is an Electric Current

An electric current is a flow of electric charge just as a water current is a flow of water or an air current is a flow of air. All metallic conductors have electric charges in them which are free to move. A drift of such charges in one direction along a conductor constitutes an electric current.

In metals these charges are electrons. As electrons carry negative charges, they move from places of lower to places of higher potential (See § 267). The motion of electrons from lower to higher potential produces the same effect as the motion of an equal positive charge from higher to lower potential. We are accustomed to look upon a current in a conductor as a flow of positive charge, though what takes place is a flow of negative charge. Remember this fact; but do not allow it to confuse you.

A current of water flowing through a pipe may be measured by the amount of water which flows past any section of the pipe in a given time, say, one second. Similarly, an electric current may be measured by the amount of charge which flows past any section of a conductor in one second. The practical unit of electric current is the *ampere*. If one coulomb of electricity moves past any section of a conductor in one second, the current is said to have a value of one **ampere**. Obviously, a current of C amperes flowing for t seconds will transport Q coulombs of electricity given by

$$Q(\text{in coulombs}) = C\ (\text{in amperes}) \times t\ (\text{in seconds})$$

In our example of water flowing through a closed, undivided system of pipes, the current at all places has the same value, i.e., the same quantity of water is transported across any transverse section of the pipe wherever we consider the section. Similarly, in the case of electricity, *an electric current will have the same value everywhere along an undivided circuit.*

Experiment 52. To set up a simple electric circuit. (Fig. 301)

Apparatus : An accumulator (B), a key (K), a rheostat (R) an ammeter (A), necessary connecting wires.

[*Descriptions*: An *accumulator* is a device which drives electricity through the circuit. It corresponds to the pump P of fig. 300. Its action is described later in § 282. In fig. 301 B represents the accumulator, the longer line standing for the gate or the *electrode* from which positive electricity flows into the circuit. The thick, shorter line is the gate or electrode through which it flows back into the accumulator. The former is called the *positive electrode* and the latter, the *negative electrode*.

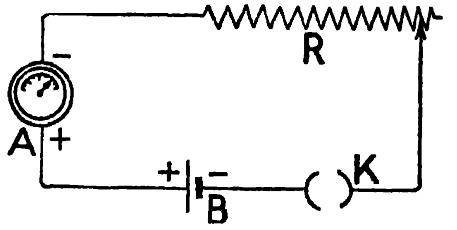


Fig. 301

An *ammeter* is a direct-reading instrument which gives in amperes the value of the current flowing through it, and hence also through the circuit.

A key or a *switch* is a very simple device which can close, (i.e., conductingly complete) the circuit or introduce an air gap (and hence a break) in it. When the circuit is broken, no current flows. So the key corresponds to the device *F* in fig. 300. Removal of the plug in a plug key (Fig. 302) breaks the circuit. Its insertion closes the circuit.

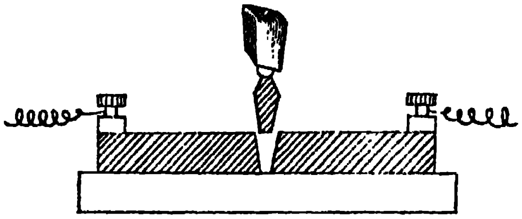


Fig. 302

A *rheostat* consists of a coil of bare wire wound in a narrow groove round a porcelain cylinder (Fig. 303). The successive turns do not touch each other. The ends of the coil are connected to two terminals, *A* and *B*. *CD* is a metal rod with a terminal at *D*. *S* is a sliding contact which moves along *CD* and brings it into metallic connection with the coil at different places along the length of the coil.

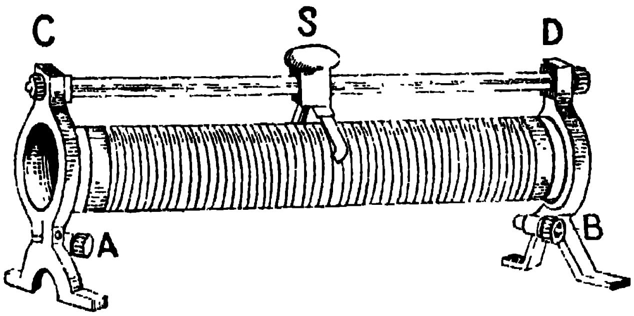


Fig. 303

If the rheostat is connected to the circuit by the terminals *A* and *B*, the whole of the coil is included in the circuit. But if the connections are at *A* and *D*, the portion of the wire included in the circuit depends on the position of *S*. The effective portion is from *A* to the point of contact of *S* on the coil as shown in fig. 304. Increasing the effective portion of a rheostat reduces the value of the current in a circuit. In this experiment start by having *S* midway between *C* and *D*.

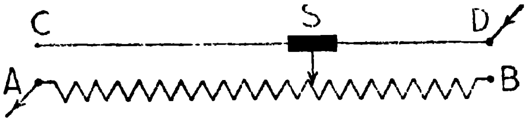


Fig. 304

The *connecting wires* are of copper wound round with cotton (or silk) for insulation. The ends must be bared of cotton, cleaned with sand paper, bent into a half loop and firmly held between the nuts of the *binding screws* with which all electrical apparatuses are provided.

In making the connections see that the terminal of the ammeter marked + (positive) is connected with the positive electrode of the accumulator.

Having completed the connections properly, and checked them insert the plug of the key in position and observe the following :

(*i*) The needle of the ammeter moves showing that a current is flowing.

(*ii*) The ammeter gives a steady reading because it measures the rate of flow of charge and not the total charge flowing in a time.

(*iii*) When the plug is removed the needle of the ammeter flies to the zero position showing that a current is no longer flowing. The removal of the plug has *broken* the metallic circuit by introducing an air gap in it. Air is a non-conductor of electricity.

(*iv*) Slide the rheostat contact *S* so as to include more of the rheostat wire in the circuit. See that the ammeter gives a lower reading ; the current in the circuit has diminished.

Move *S* in the opposite direction. The current will increase.

274. Three Effects of an Electric Current

An electric current is known to produce three kinds of effects, namely, (*i*) *magnetic,* (*ii*) *thermal* and (*iii*) *chemical* or *electrolytic.* The fact that an electric current always produces a magnetic field around it is known as its magnetic effect. When an electric current passes through a conductor, it generates heat in the conductor. This is known as the heating effect. An electric current passing through conducting solutions (such as solutions of acids, salts or bases in water), brings about chemical decomposition of the molecules in solution. This is the chemical or electrolytic effect of a current.

Some simple experiments to demonstrate the effects are described below.

Experiment 53. To demonstrate the magnetic effect of a current.

Apparatus: As in Experiment 52, a connecting wire about a metre long, a magnetic needle (preferably long).

Set up the circuit including the metre-long wire in it and adjust the rheostat so that the current is about one ampere. Now, observe the following :

(*i*) Hold the long wire parallel to the needle and *above* it. The needle will be deflected. Observe the direction in which the north pole moves.

(*ii*) Bring the wire nearer the needle. The deflection increases.

(*iii*) Increase the current. The deflection increases.

(*iv*) Now hold the wire *below* the needle and repeat the observations. The deflection of the north pole will be in the opposite direction.

(*v*) Interchange the connections at the two poles of the accumulator. This reverses the direction of flow of current in the circuit. (Note that this will also necessitate interchanging connections at the ammeter unless it is a centre-zero instrument. Reversal of the current relative to the needle can be more easily done by turning the wire or inverting it).

Repeat observations (*i*) to (*iv*) and note that in each case the deflection now is in the opposite direction.

(*vi*) Hold the wire vertically (*a*) beyond the N-pole, (*b*) beyond the S-pole. Observe the deflection of the N-Pole in each case, and also when the current is reversed in either position.

Complete the following table, remembering that the conventional electric current flows in the external circuit from the positive pole to the negative pole of the accumulator. (Poles are the external terminals of the electrodes.)

Position of current relative to needle	*Direction of current flow*	*Direction in which N-pole is deflected (Say if East or West)*
Above	South to North	—
	North to South	—
Below	South to North	—
	North to South	—
Beyond the N-pole	Vertically upwards	—
	Vertically downwards	—
Beyond the S-pole	Vertically upwards	—
	Vertically downwards	—

From your observations confirm the following rules :

(i) **Oersted's* rule.** If the right palm is so placed that the current lies between it and the needle (Fig. 305), and the fingers point in the direction of the current, the out-stretched thumb will show the direction in which the north pole is deflected.

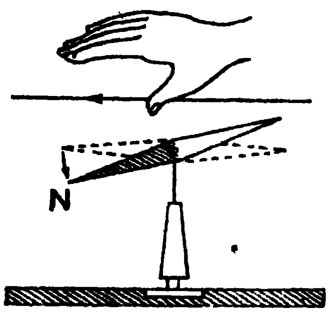


Fig. 305

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* The fact that an electric current produces a magnetic field around it was accidentally discovered in 1819 by Hans Christian Oersted of the Copenhagen University while conducting an experiment before his class. He found that a wire carrying a current deflected a magnetic needle when held near it. When the current was reversed, the deflection was in the opposite direction.

(ii) **Ampere's rule.** If we imagine a man swimming along the conductor in the direction of the current with his face towards the needle (Fig. 306) the north pole of the needle will be deflected towards his left hand.

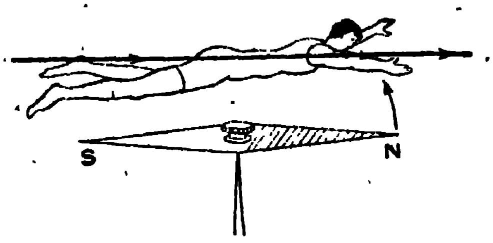

Fig. 306

Other magnetic effects of currents are discussed in Chapter 35.

Experiment 54. To demonstrate the heating effect of a current.

Apparatus: As in Experiment 52, but with two accumulators, a piece of nichrome wire about a metre long (strectched between two stands, or wrapped round an asbestos sheet) with binding screws attached to the ends.

Set up the circuit including the nichrome wire in it. Use both accumulators, connecting the positive of one to the negative of the other, and the circuit to the free poles of the two. Let the whole of the rheostat wire be included in the circuit at first.

Gradually increase the current by adjusting the rheostat The nichrome wire will feel warmer to the touch. Be careful it does not singe your finger. As you go on slowly increasing the current the nichrome wire will begin to glow at some stage. Further increase in current will make it glow brighter. It may eventually melt. Do not, however, allow the current to rise above one ampere lest your rheostat should be damaged. You will see that as the current increases the rheostat wire is appreciably heated. (A rheostat which can stand currents of over one ampere should be used).

Electric lights are due to the heating effect of currents. The fine tungsten wires in the bulbs are heated to such high temperatures by the currents flowing through them, that they emit bright light. You will see that the bulbs are warm to the touch when glowing.

Electric stoves, heaters and irons are other examples of the application of the heating effect of currents. The effect has been further discussed in Chapter 36.

Experiment 55. To demonstrate the chemical action of a current.

Apparatus: As in Experiment 54 but without the nichrome wire; a glass cell; two copper strips of the shape of an elongated *S*; some sulphuric acid,

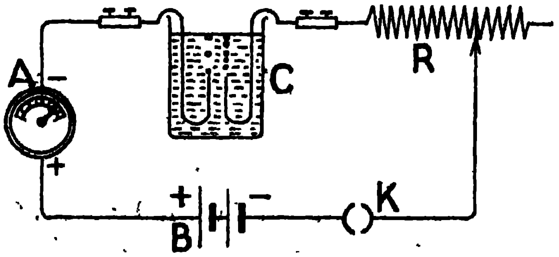

Fig. 307

Set up the circuit with the cell included in it as in fig. 307. Suspend the copper strips from the edge of the glass cell. Their outer ends should have some attachment for holding connecting wires.

No current passes through the circuit as there is a large air gap in the glass cell. Pour water into the cell. The current still does not pass. Water is a non-conductor of electricity.

Pour a little sulphuric acid into the water and stir. The current will now pass. Add more acid until the current increases no further. Now adjust the rheostat so that the current is about an ampere.

You will find that bubbles of gas are rising from the copper strips. These gases can be collected over water in inverted test-tubes filled with water (Fig. 308). For this purpose the strips may be varnished except at the ends. On testing you can prove that one of the gases is oxygen and the other hydrogen. When the current is stopped the evolution of gas ceases.

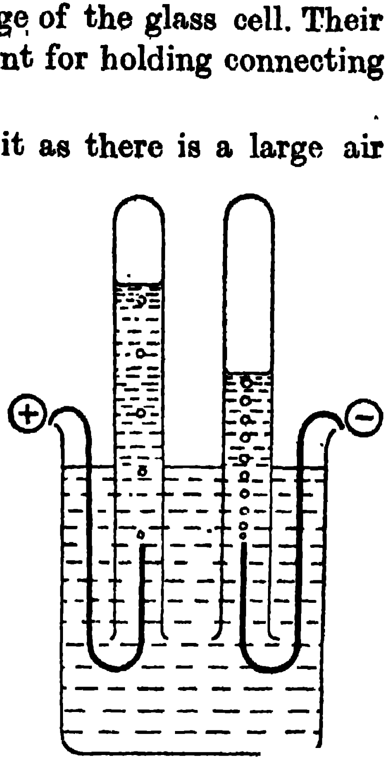

Fig. 308

The passage of the current through the solution has clearly brought about some chemical action—here decomposition of water into hydrogen and oxygen. It is an example of the chemical effect of currents. With copper and sulphuric acid, as used above, a blue coloration will soon appear. This is due to copper getting into solution.

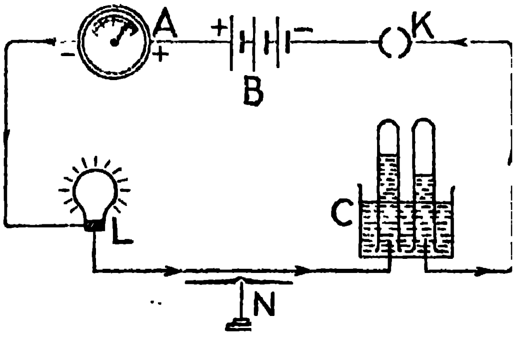

Fig. 309

Other examples and applications of chemical effects of currents have been discussed in Chapter 37.

The three kinds of effects can be demonstrated together in the same circuit as in fig. 309. *L* may be the bulb of a torch. Its glow demonstrates the heating effect. *N* is a magnetic needle. A portion of the circuit wire runs parallel to *N*. *C* is a glass cell as described in Experiment 55. *A* is an ammeter and *K*, a key. *B* may be a motor car battery or a group of two or three accumulators,

275. Physiological Effects of an Electric Current

Most of you who use electricity at home might have, some time or other, received an electric 'shock' while handling a faulty electrical equipment, such as a defective bulb, fan, heater or fuse etc. The 'shock' is the effect of the passage of an electric current through your body to the earth.

An electric current passing through the body will affect both sensory and motor nerves, and thus cause sensations of pain, involuntary movements and twitchings. Strong shocks, due to large currents, may cause severe burning and death by heart failure.

The degree of shock depends on the current producing it. The current in its turn is determined by two factors: (*i*) the 'voltage' available to drive the current, and (*ii*) the intimateness of the contact between the body and the electrical equipment on the one hand and the body and the earth on the other. More often the current passes through the arm, body, and legs to the earth. If the hands and legs are moist the current will be stronger. Higher voltages produce stronger currents.

To protect yourself from a shock stand on a dry wooden stool or board while handling electrical equipment connected with supply mains. Potential differences of 20 or 30 volts hardly produce any sensible shock.

Exercises

1. Point out the similarities between a current of water and a current of electricity. What part do the following devices play in an electric circuit: (*a*) a source of e.m.f; (*b*) a key; (*c*) a rheostat; (*d*) an ammeter?

2. State Oersted's rule and Ampere's rule relating to the deflection of the north pole of a magnetic needle by an electric current.

A current flowing north to south is held below a magnetic needle. Apply any of the above rules to find the direction of deflection of the south-pole. Draw a diagram to show application of the rule in the given case.

3. How would you arrange an electric circuit to demonstrate all the three principal effects of an electric current? Illustrate with a neat diagram.

4. What is an electric circuit? How should connecting wires be used in a circuit to connect the appliances included in the circuit? When an ammeter and a key are included in the circuit, is it necessary that they should occupy certain fixed positions relative to the source of e.m.f.?

CHAPTER 33

ELECTROMOTIVE FORCE AND VOLTAIC CELLS

276. What keeps an Electric Current flowing

Any motion or any change, as you know, requires an agency to bring it about, and also needs an expenditure of energy. All the time the motion or the change continues, the agency must be acting and spending energy. The flow of an electric current is no exception. Some agency must be there to drive the current, and it must have the necessary energy at its disposal to expend in the action.

By the term **electromotive force** (abbreviated e.m.f.) we mean an agency which, when present in a conducting circuit, can drive an electric current. Now, electric charges do not move unless there is a potential difference (See § 267). The action of the source of e.m.f. consists in setting up and maintaining a potential difference. The e.m.f. of a source is measured by the potential difference it sets up. It is expressed in *volts* (See 267).

Energy of the source of e.m.f. The source of e.m.f. must have the necessary energy at its disposal to drive the current. Where does this energy come from? In the various types of voltaic cells (discussed in subsequent sections) the energy comes from the chemical actions taking place within the cell. Here chemical energy is converted into electrical energy. In other kinds of sources of e.m.f., other forms of energy are converted into electrical energy, e.g., mechanical energy in a dynamo, heat energy in a thermo-couple and light energy in a photovoltaic cell.

Remember that a source of e.m.f. is a device which converts energy into an electrical form.

Conversion of electrical energy into other forms. Energy cannot be destroyedjust as it cannot be created. What does then happen to the electric energy a source of e.m.f. supplies? A part of it, generally a small part, is expended in creating the magnetic field around the current. If there is nothing else for the current to do, it will expend the rest of the energy in developing heat in the conductors through which it passes. Some times the current has other things to do as well, such as turning the motor of a fan or machine or causing electrolysis. In the former case electrical energy is also converted into mechanical energy; in the latter case, into chemical energy. The development of heat is, however, always there.

277. The Simple Voltaic Cell

Experiment 56. To construct a simple voltaic cell and study its action.

Apparatus: Amalgamated zinc rod[1] with binding screw attached, copper plate with binding screw, dilute sulphuric acid in a porcelain pot, copper wires, magnetic needle, ammeter.

Observations: (*i*) Place the zinc rod in the acid. There is no visible action. Place the copper plate also in the acid taking care that it does not touch the zinc. No action occurs.

(*ii*) Connect the two metals by a wire outside the liquid. You will find that zinc begins to dissolve and bubbles of hydrogen appear on the copper plate.

(*iii*) Hold the connecting wire above the magnetic needle and parallel to it. The needle will be deflected.

(*iv*) Use a fine wire to connect the metals. It will be heated.

[To get the effects at (*iii*) and (*iv*) clearly, it will be necessary to brush away the gas bubbles sticking to copper and zinc.]

From (*iii*) and (*iv*) you may conclude that an electric current is flowing through the wire.

Action of the cell. The above combination of zinc, acid and copper is known as a simple voltaic[2] cell.

The zinc dissolves as the current flows. The chemical energy released in the action is converted into electrical energy. One may therefore, say that *the 'fuel' in a simple voltaic cell is zinc.*

The electromotive force is set up by the chemical action which takes place at the surfaces of contact between the metals and the acid. It drives positive electricity to the copper and negative electricity to the zinc. The presence of a positive charge on the copper and a negative charge on the zinc can be established experimentally. The potential difference between them can also be measured. Its value when the cell is not driving any current is the measure of the e.m.f. In this case it is nearly one volt.

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[1] An amalgam means a solution of one or more metals in mercury. To form the amalgam, the zinc should be cleaned in acid, and a few drops of mercury worked into its surface until a smooth, even, somewhat plastic surface is obtained.

[2] It is named after Alessandro Volta, an Italian scientist, who in 1800 A.D. set up the device for the first time.

When a current is passing, electrons move along the wire from the zinc to the copper and neutralize the positive charge there. The e.m.f. keeps the metals continuously supplied with electricity so long as there is zinc and acid in the cell to provide the chemical action.

Conventionally, we treat the current as a flow of positive charge. Hence the direction of flow is taken as from copper to zinc externally and from zinc to copper within the cell.

The copper plate within the cell is called the positive electrode, and its outer terminal, the *positive pole* of the cell. The corresponding parts of zinc are the negative electrode and the *negative pole*. The acid is called the *electrolyte*.

Other simple cells. A simple cell can be made with a variety of materials. Stick a piece of copper wire and a zinc wire into a lemon or an orange. The wires become the electrodes of a simple cell of which the electrolyte is citric acid. Test by touching your tongue with the wires. The sensation you get on the tongue is due to the passage of a small electric current.

Put a small piece of blotting paper soaked in salt solution between two coins of different metals. This combination is also a simple cell.

Ordinarily, any acid can serve as the electrolyte. The positive electrode may be carbon, or a metal like silver, gold, platinum etc., which does not readily dissolve. The negative electrode should be a material like zinc, lead, aluminium etc., which dissolves more readily. The e.m.f. will depend on the nature of the electrodes and the electrolyte.

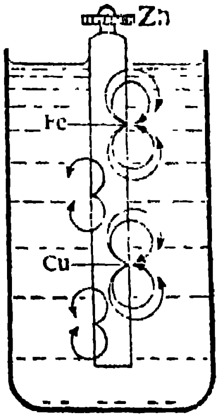

Fig. 310

278. Defects of the Simple Cell

The simple cell has two main defects known as (*i*) local action and (*ii*) polarization.

(*i*) **Local action.** Common zinc contains impurities, such as carbon, iron, lead, arsenic etc. The particle of impurity acts as the positive pole of a tiny local cell on the surface of zinc. Electricity flows in a short local circuit from the impurity to the zinc and back through the electrolyte to the particle (Fig. 310). This action is known as *local action*. It results in a wastage of the electrode, usually the zinc, and does not produce any flow of electricity in the external circuit.

Local action is prevented by amalgamating the zinc with mercury. The mercury dissolves the zinc and the amalgam covers up the impurities. As the zinc is consumed the impurities fall to the bottom of the cell.

The electrodes should be removed from the acid when the cell is not in use.

(*ii*) **Polarization.** When the simple cell supplies a current, bubbles of hydrogen gas collect on the copper electrode. It makes the electrode behave like hydrogen rather than like copper, and reduces the e.m.f. of the cell. This effect is known as polarization. More generally we can say that *polarization* is the change in the nature of an electrode brought about by the chemical action going on within the cell.

Since hydrogen gas is a non-conductor of electricity the deposition of hydrogen on copper in a simple cell retards the flow of electricity in the circuit and quickly reduces the value of the current.

Polarization is minimized by surrounding the positive electrode with an oxidizing agent, usually manganese dioxide (MnO_2) or potassium bichromate ($K_2Cr_2O_7$). It is called a *depolarizer* and its function is to combine with the offending hydrogen as it forms. In all types of cells used in practice, devices are employed to prevent or reduce polarization.

279. The Leclanche Cell

The Leclanche cell, designed by Georges Leclanche, has a negative electrode of zinc (*Zn*; Fig. 311), and a positive electrode (*C*) of carbon. The latter is packed round with a mixture (*A*) of manganese dioxide and powdered carbon, and enclosed in a porous pot (*P*). The electrolyte (*S*), contained in a glass vessel, is a solution of ammonium chloride and surrounds the porous pot. The solution diffuses through the pot and impregnates the mixture around the carbon rod.

When a current is drawn from the cell, zinc goes into solution, and ammonia and hydrogen are liberated at the positive electrode. The ammonia dissolves in the ammonium chloride solution and the hydrogen is oxidized to water by the manganese dioxide which acts as depolarizer.

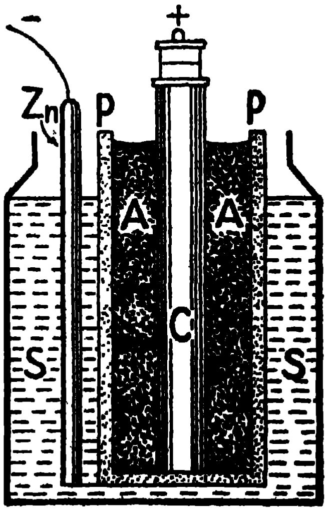

Fig. 311

The oxidation takes place slowly. So when a current is taken for some time, hydrogen slowly accumulates and the cell polarizes. Left to itself it soon depolarizes and, after a rest, is ready for use again.

The e.m.f. is 1·5 volts. The cell is well suited for intermittent operation. It is used with electric bells and flash lights, in which the current is taken only for short periods.

The Leclanche cell has good many advantages. It can be left for months without attention, has a low running cost and gives little trouble. It is occasionally necessary to add a little water to make good the loss due to evaporation. A general cleaning and replacement of ammonium chloride solution may be done once a year.

280. The Dry Cell

The 'dry cell' is really a form of Leclanche cell in which the fluid has been replaced by a damp mixture of the necessary ingredients. The cell would not function if it were dry. The term 'dry cell' is, therefore, a misnomer. One may say it is 'non-spillable'.

The cell usually has a zinc can (Fig. 312) which serves as the container and is also the negative electrode. It is lined with blotting paper saturated with a solution of ammonium chloride and zinc chloride, or with a paste composed of plaster of Paris, flour, zinc chloride, ammonium chloride and water. A carbon rod, which forms the positive electrode, is at the centre of the zinc can. The space between the rod and the lining on the zinc is filled with a paste made of granulated carbon and manganese dioxide soaked in a solution of ammonium chloride and zinc chloride. Zinc chloride is highly hygroscopic and keeps the cell from becoming too dry. The contents are sealed with pitch, through which a vent may be left to permit gases to escape.

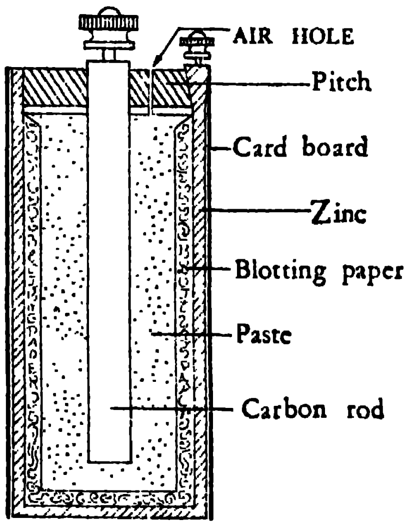

Fig. 312

The e.m.f. of the cell is 1·5 volts. Millions of dry cells are used daily in flash lights, radio sets and in operating electric bells, telephones, signal devices etc. It has all the advantages of a Leclanche cell and is non-spillable in addition.

281. The Daniell Cell

Another useful form of cell is the Daniell cell. It consists of a positive electrode of copper in a saturated solution of copper sulphate (Fig. 313), and a negative electrode of amalgamated zinc in a 5% solution of zinc sulphate containing a little sulphuric acid. The two electrolytes are separated by a porous pot of unglazed porcelain, which prevents the solutions from mixing but allows the flow of current. Copper sulphate crystals are often kept on a perforated tray to maintain the strength of the solution.

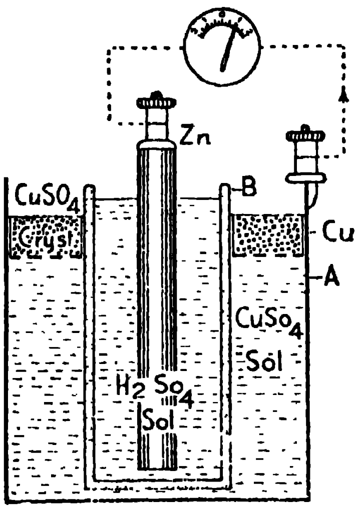

Fig. 313

During operation of the cell zinc dissolves. The hydrogen, instead of forming on the positive plate, displaces copper from the copper sulphate. This copper is deposited on the positive plate. Hence there is no polarization. The cell has a steady e.m.f. of about 1·08 volts. It is used in the operation of railway signal circuits where a continuous small current is necessary. Its great disadvantage is that it must be dismantled when not in use. Otherwise the solutions mix up by diffusion through the pores of the pot, and the zinc dissolves gradually.

282. Accumulators

The cells described so far are known as **primary cells.** They convert chemical energy into electrical energy, and when their store of chemical energy is exhausted they cease to function unless more chemicals are added.

There is another kind of cells known variously as **accumulators, storage cells** or **secondary cells.** In them electrical energy is stored as chemical energy by a process known as 'charging'. When 'charged', the cell is capable of supplying electrical energy by chemical action in the same way as a primary cell. There are two principal types, viz., (*i*) the *lead accumulator* or *lead-acid cell,* and (*ii*) the *Edison cell,* also called *nickel-iron-alkali cell* or *NiFe cell.*

The lead accumulator. In a commercial cell which has been charged and is ready to deliver a current, the negative electrode consists of a group of parallel plates of pure spongy lead. Interposed between them lies a group of lead grids packed with porous packets of lead peroxide (PbO_2) which form the positive electrode. The two sets of plates are kept apart by glass, rubber or wooden separators. Positives and negatives alternate with each other. All the positive plates are connected to the same terminal, and all the negatives to another. The plates are immersed in an electrolyte of about 20% sulphuric acid in a glass container.

While 'discharging', i.e., delivering current, both plates (+ve and −ve) become coated gradually with lead sulphate and their colour changes to grey. The electrolyte becomes less dense with discharge. The e.m.f. his a steady value of 2 volts. The cell can deliver large currents. It is a very useful and common appliance in an electrical laboratory.

The process of storing chemical energy in the cell by electrical means is known as 'charging'. During 'charge' an electric current is sent through it in a direction opposite to that in which it discharges. This brings about chemical changes within the cell, restoring it to the condition in which it can deliver electrical energy.

The nickel-iron-alkali cell. In this cell the negative electrode consists of a mixture of finely divided iron and iron oxide (FeO) contained in perforated steel pockets. The positive electrode consists of alternate layers of nickelic hydroxide and nickel flakes packed into perforated steel boxes. The electrolyte is a 21% solution of potassium hydroxide (KOH).

The cell has an average e.m.f. of about 1·2 volts. It is more expensive than the lead type, but is more robust and stands rough handling better.

283. Some Facts to remember

The following facts may be remembered with profit :

(*i*) The size of a cell does not determine its e.m.f.

(*ii*) The e.m.f. of a cell is determined by the chemical actions in the cell, i.e., by the nature of the electrodes and the electrolyte.

(*iii*) The total amount of charge a cell can deliver is determined by the amount of reacting chemicals in the cell.

(*iv*) The surfaces of contact between the electrodes and the electrolyte are the seat of the e.m.f.

(*v*) The e.m.f. keeps the two electrodes supplied with two kinds of electricity.

Exercises

1. What is an electric current ? What is meant by the term *electromotive force* ? How do we get the energy required to drive an electric current in a circuit ? (Write the answer to each in only one sentence.)

Give an example each of conversion of electrical energy into some other form, including at least one example not given in your book.

2. Describe a simple voltaic cell and state how it acts. On what factor does its e.m.f. depend ?

3. What are the defects of a simple cell as a source of e.m.f. ? How have these defects been removed in the case of the Daniell cell ?

Draw a sketch of a *dry cell* and label the parts. Why is it called *dry* ? What is the kind of work for which it is suited ? What is its e.m.f. ?

5. What are accumulators ? In what essential respect does an accumulator differ from a cell like the Leclanche cell ?

Briefly describe any accumulator you are familiar with, stating what electrodes and electrolyte it has, and its e.m.f.

6. Answer the following :

(*a*) Why is zinc considered the fuel of a cell ?
(*b*) What determines the e.m.f. of a cell ?
(*c*) Where is the seat of the e.m.f. in a cell ?
(*d*) Why is the zinc rod in a Leclanche cell amalgamated ?
(*e*) What part is played by (*i*) manganese dioxide, (*ii*) carbon granules, in a dry cell ?

CHAPTER 34

RESISTANCE AND OHM'S LAW

284. Resistance

We have already seen that there is similarity in many respects between an electric current and a current of water. Here we shall have a further example. When water flows through a pipe under pressure, the pressure at the outlet end is less than that at the inlet end. The pipe offers a resistance to the flow of water. A part of the pressure (i.e., the head of water) is used up in overcoming this resistance. The longer or the narrower the pipe, the greater is its resistance.

A conductor will similarly offer a resistance to the flow of current through it. In connexion with Experiments 52 to 55 we have already seen that the current is smaller when a longer portion of the rheostat wire is included in the circuit. In these experiments the potential difference applied to the circuit was maintained constant by the e.m.f. of the accumulator. When, therefore, we increased the resistance of the circuit by increasing the length of the wire through which electricity had to flow, a smaller current was obtained.

285. Ohm's Law

The relation between potential difference and the resulting electric current was investigated by G.S. Ohm, a German physicist. His result, announced in 1826 as Ohm's law, is one of the greatest importance and widest application in electricity. It may be stated as follows :

Ohm's law: The current flowing between any two points of a conductor is proportional to the potential difference between them, provided that the temperature and other physical conditions of the conductor do not change.

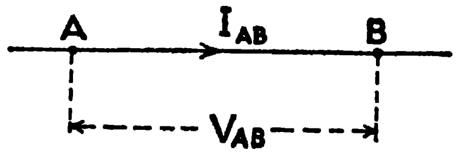

Fig. 314

In fig. 314, let A and B be two points between which there is a steady potential difference, V_{AB}, causing a steady current, I_{AB} to flow. Since, according to Ohm's law, V_{AB} and I_{AB} are proportional, the ratio of the former to the latter is a constant.

$$\therefore \quad \frac{V_{AB}}{I_{AB}} = \text{Constant} = R_{AB} \text{ (say)}. \qquad (285{\cdot}1)$$

This ratio, R_{AB}, is called the *resistance* of the conductor between the points A and B. Its value depends on the material, dimensions and temperature of the conductor, as we shall see later.

Dropping subscripts, we may write Eq. 285·1 as

$$I = \frac{V}{R}, \text{ or } R = \frac{V}{I} \text{ or } V = IR \quad (285{\cdot}2)$$

Any of these equations may be taken as the mathematical form of Ohm's law. In applying it the correct meanings, as we have stated above, must be assigned to V, I and R. Students, unfortunately, are too prone to neglect this simple advice, with the result that even at a very late stage incorrect applications of Ohm's law are made.

Ohm's law, as stated above, applies to an undivided part of a circuit, i.e., to points of a conductor between which the current neither divides nor augments.

Ohm—the unit of resistance. The relation $R = V/I$ is used in defining the unit of resistance. When a potential difference of one volt drives a current of one ampere through a conductor, it is said to have a resistance of one **ohm.**

$$\textbf{Ohms} = \frac{\textbf{Volts}}{\textbf{Amperes}}$$

To realize the ohm in practice the following definition was accepted by international agreement. The **international ohm** is the resistance offered to an unvarying electric current by a column of mercury at 0°C, 14·4521 gm in mass, of constant cross-section and of length 106·300 cm. The international ohm represents the 'true' ohm very closely.

Other International Units. For convenience the ampere and the volt have also been defined by international agreement in terms of easily realizable quantities. Thus

The **international ampere** is that unvarying current which deposits silver at the rate of ·001118 gm per sec when passed through a solution of silver nitrate in water.

The **international volt** is that potential difference which produces a current of one international ampere when applied across a resistance of one international ohm.

286. Verification of Ohm's Law

Using an ammeter to measure currents in amperes and a voltmeter to measure potential differences in volts, we can verify Ohm's law as follows :

Experiment 57. *To verify Ohm's law.*

Apparatus. An ammeter, a voltmeter, a battery of two or more accumulators, a rheostat, a manganin or nichrome wire wound round an asbestos sheet and provided with terminals for binding two wires, connecting wires, a key.

Set up the circuit as shown in fig. 315. *B* is the battery; *A*, the ammeter; *r*, the rheostat, *R*, the manganin (or the nichrome) wire; *V*, the voltmeter connected to the two ends *P* and *Q* of *R*. The voltmeter gives the potential difference in volts across the points to which its terminals are connected—here the ends of *R*. See that the positive terminals of the voltmeter and ammeter are connected to the positive pole of the battery.

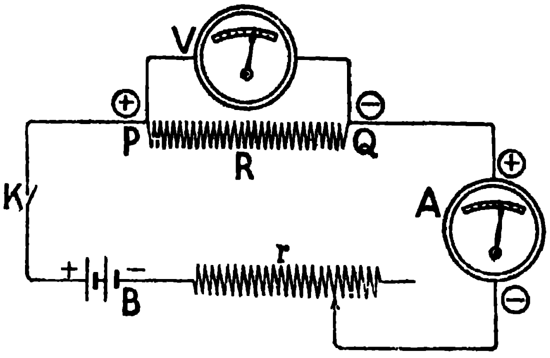

Fig. 315

Switch on the current and note the readings of the voltmeter and ammeter. Vary the current with the help of the rheostat and note down a number of corresponding readings of the voltage and the current. Tabulate the results as follows :

Serial number	*Potential difference in volts (V)*	*Current in amperes (I)*	$\frac{V}{I}$ = *R ohms*	*Mean R*
				....ohms

If the currents were not large enough to heat *R* appreciably, or the room temperature did not change much during the experiment it will be seen that the values in column four are constant within limits of experimental error. This constancy verifies Ohm's law. The mean of the values in column four gives the resistance of the wire in ohms.

287. Factors affecting Resistance

We can arrive at some useful and important results if in Experiment 57 we replace the conductor *R* successively by

(*i*) different lengths (*l*) of wires of the same material and cross-section,

(*ii*) equal lengths of wires of the same material but different cross-sectional areas (*A*), and

(*iii*) equal lengths of wires of the same cross-sectional area, but made of different metals, If the temperature remains constant we shall find that

(*a*) For wires of the same material and cross-section, the resistance is directly proportional to the length, or

$$R \propto l, \text{ if } A \text{ is constant.}$$

(*b*) For wires of the same material and length, the resistance is inversely proportional to the cross-section, or

$$R \propto \frac{1}{A}, \text{ if } l \text{ is constant.}$$

(*c*) For wires of the same length and cross-section the resistance depends on the material of the wire.

Combining (*a*) and (*b*) we may write

$$R = \rho \frac{l}{A}$$

The value of ρ depends on temperature and the nature of the material. It is called the *resistivity* or the *specific resistance* of the material, and is defined as follows :

The **resistivity** or **specific resistance** of a material is the resistance of one centimetre length of the material when the area of cross-section is one square centimetre. In other words, it is the resistance between the opposite faces of a cube of the material of side one centimetre.

Unit of resistivity. To find a unit in which resistivity may be expressed let us remember that

$$R \text{ (in ohms)} = \rho \times \frac{l \text{ (in cm)}}{A \text{ (in cm}^2)}$$

$$\therefore \quad \rho = \frac{R \text{ (ohm)} \times A \text{ (cm}^2)}{l \text{ (cm)}} = \frac{RA}{l} \text{ ohm} \times \text{cm.}$$

In engineering it is customary to measure l in feet and A in *circular mils.* (A mil is one-thousandth of an inch.) ρ is the resistance of a mil-foot of the material, i.e., of a wire 1 ft long and ·001 inch in diameter.

Silver is the best conductor and copper a close second. The following table gives the values of the resistivities of a number of metals and alloys.

Metals	*Specific resistance*	*Alloys*	*Specific resistance*
	$\times 10^{-6}$ ohm.cm		$\times 10^{-6}$ ohm. cm
Silver	1·7	Brass	6 to 9
Copper	1.8	German Silver	16 to 40
Aluminium	3.2	Manganin	42 to 45
Iron	12 to 20	Eureka (Constantan)	49
Lead	20.8		
Mercury	95.8	Nichrome	110

Resistance and temperature. The resistivity of a substance varies with the temperature, and increases as the temperature rises. The change is about 0·4% per °C in the case of most metals. For alloys it is much less. The resistance of manganin changes very little with temperature.

Resistance and other physical conditions. In general resistance is little affected by physical conditions other than temperature. There are some special cases, however. The resistance of selenium (a semi-metallic element) is affected by the intensity of the light falling on it. This has made it posible to use selenium cells for automatic street-lighting. Resistance diminshes when the intensity of light increases.

In bismuth the resistance is affected by a magnetic field. The change in resistance of a bismuth spiral in a magnetic field has been utilized for measuring the strength of the field. In stronger fields the resistance is higher.

288. Resistance Boxes

A resistance box consists of a number of resistance coils so mounted in a box that a current can be sent through one or more of the coils as desired. The ends of a coil are connected to two brass blocks mounted on a sheet of ebonite or vulcanite. The blocks (*M*; fig. 316) are separated from each other. A coil may be short-circuited by putting a conical brass plug (*P*) into the conical hole (*G*) between the blocks. The coil (*C*) is folded on itself and wound round a bobbin (*B*). (Such winding. called a *non-inductive winding*, minimizes the magnetic field which the coil produces).

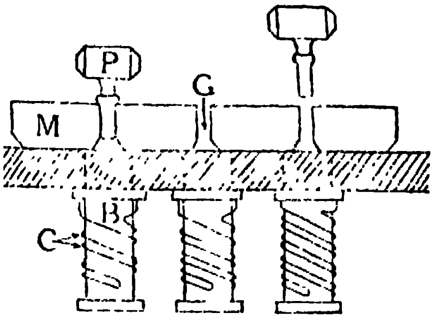

Fig. 316

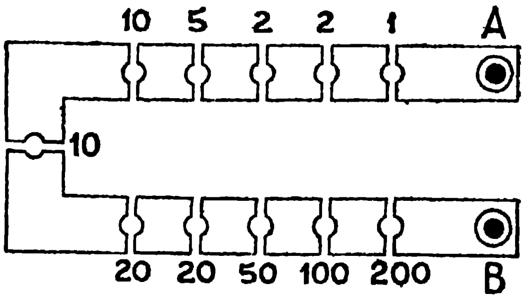

Fig. 317

A number of such coils are arranged in a box, as may be understood from figs. 316 and 317. The numbers in fig. 317 indicate in ohms the value of the resistance of the coil lying between the two neighbouring brass blocks. *A* and *B* are terminals where to fasten connecting wires of the circuit. When a plug is removed the corresponding resistance becomes operative. With the plug in position the current passes through the plug and not through the wire which the plug short-circuits.

289. Resistances in Series and in Parallel

Resistances are said to be joined **in series** when they are so arranged that the same current passes through all of them. To arrange a number of resistances in series, one end of the first is connected to one end of the second; the other end of the second is connected to one end of the third and so on (Fig. 318). The total resistance of such a set of conductors is equal to the sum of the resistances of the individual conductors. In fig. 318 the total resistance $R = R_1 + R_2 + R_3$.

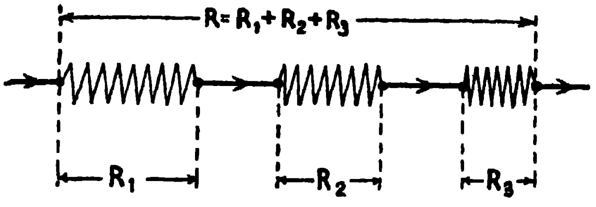

Fig. 318

When two or more resistances provide alternative paths for the current, a part of the current flowing through each, the resistances are said to be joined **in parallel.** Fig. 319 shows three resistances R_1, R_2 and R_3 joined in parallel. To proceed from A to B a current in the main circuit divides along these three paths. Let I_1, I_2 and I_3 be the currents in the three resistances. If V is the potential difference between A and B, we have, from Ohm's law,

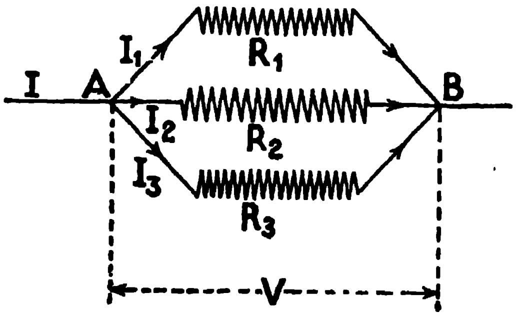

Fig. 319

$$V = I_1R_1 = I_2R_2 = I_3R_3$$

or $$I_1 = \frac{V}{R_1}, I_2 = \frac{V}{R_2}, I_3 = \frac{V}{R_3}$$

If R is the effective or equivalent resistance between A and B, we have, again from Ohm's law,

$$R = \frac{V}{I} \text{ or } I = \frac{V}{R}$$

where I is the total curent between A and B,

But $I = I_1 + I_2 + I_3$.

$$\therefore \quad \frac{V}{R} = \frac{V}{R_1} + \frac{V}{R_2} + \frac{V}{R_3}$$

or $$\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3}$$

In words we may say that the reciprocal of the equivalent resistance of a number of conductors in parallel is equal to the sum of the reciprocals of the component resistances.

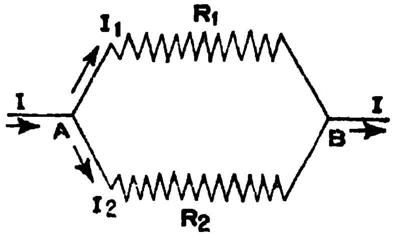

Fig. 320

Two resistances in parallel. It is easy to show in this case that

$$\frac{1}{R}=\frac{1}{R_1}+\frac{1}{R_2} \text{ or } R=\frac{R_1R_2}{R_1+R_2}$$

and since $I_1R_1=I_2R_2$,

$$\frac{I_1}{I_2}=\frac{R_2}{R_1}$$

showing that currents in the branches are inversely proportional to their resistances.

Further, since $I=I_1+I_2$, we shall have

$$I_1=I\times\frac{R_2}{R_1+R_2} \text{ and } I_2=I\times\frac{R_1}{R_1+R_2}$$

In the case of two resistances, 1 ohm and 2 ohms, in parallel the equivalant resistance is 2/3 ohm. Note that this value is lower than either of the component resistances. This is a general result. We shall do well to remember that *when two or more resistances are joined in parallel, the equivalent resistance is lower than the lowest of the component resistances.*

Shunt. When a resistance is connected in parallel across another, the former is said to act as a *shunt* to the latter. If the shunt has a lower resistance than the other, more current will flow through the shunt.

290. Cells in Series and in Parallel

A set of connected cells is called a *battery.*

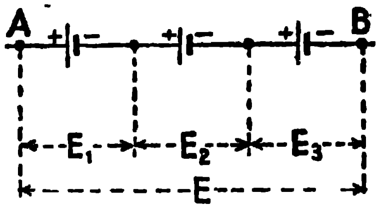

Fig. 321

Cells are said to be joined *in series* if the positive of one is joined to the negative of the next (Fig. 321). If all the cells of a battery are connected in series its total e.m.f. is the sum of the e.m.f's of the component cells.

When a number of cells having the same e.m.f are so arranged that their positive poles are connected together, and so also the negatives (Fig. 322) they are said to be joined *in parallel.* Such an arrangement is equivalent to only one cell of the same e.m.f.

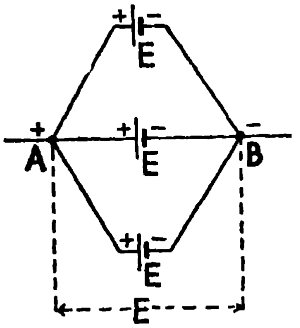

Fig. 322

When we want to draw a stronger current than what a cell can supply, we should put two or more of them in parallel. Cells are joined in series when we want a higher voltage. Wireless high tension batteries consist of about 80 small dry cells joined in series, and this provides an e.m.f. of 120 volts.

291. Internal Resistance of a Cell

The current flowing in a circuit also flows through the electrolyte in the cell which drives the current. Within the cell the current flows from the negative electrode to the positive electrode. The resistance of the liquid between the electrodes is called the *internal resistance* of the cell.

The magnitude of the internal resistance depends on the nature of the electrolyte, the distance apart of the electrodes, their size and other factors. Some approximate values are given below.

Leclanche cell	1 to 5 ohms.
Dry cell	0·1 to 0·5 ohms.
Daniell cell	0·3 to 15 ohms.
Lead accumulator (medium size)	about 0·01 ohm.
NiFe accumulator (medium size)	,,

When cells are joined in series their internal resistances should be added together to get the effective internal resistance of the battery.

If n identical cells are joined in parallel, their internal resistances are also in parallel. Hence the effective internal resistance will will be r/n where r is the internal resistance of each cell.

292. Lost Volts or Inner Drop of Potential

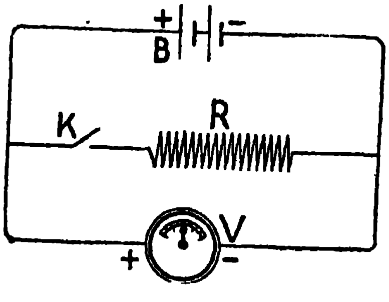

Fig. 323

Set up a circuit as shown in fig. 323. The battery B may consist of two Leclanche cells in series. R may be a resistance of about 10 ohms. V is a voltmeter and K, a key.

When the key is off, a very small

current passes through the voltmeter. Its reading may be taken to give the e.m.f. of the battery. Note the value (E, say).

Close the key. A current flows through R also. The voltmeter gives a lower reading (E', say).

The reason for this drop is not difficult to find. The e.m.f. of the cell drives the current through the external resistance as well as the internal resistance. The first reading (E) of the voltmeter gave the e.m.f. of the cell. The second reading (E') gave the potential difference between the ends of the external resistance. $E-E'$ is the potential difference required to drive the current through the cell. This difference is called the *lost volts* or the *inner drop of potential*. Its value will be Ir where r is the internal resistance of the cell and I the current through it.

We find a similar drop occurs when a current is used at a considerable distance from a generator. The voltage at the receiving end of the line is always less than the generator voltage. The voltage drop in the line is the product of the current I and the resistance R of the line, and is known as IR drop. For house supply it should not exceed 2% of the generator voltage.

293. Ohm's Law applied to Simple Circuits

A circuit in which the same current flows through all the resistances may be called a simple circuit. The cells as well as the external resistances are obivously joined in series in a simple circuit.

Let E be the e.m.f. of the battery. R the total resistance of the external conductors, r the internal resistance of the battery and I the current in the circuit. A potential difference of Ir will be required to drive the current through the cell. The balance, i.e., $E-Ir$, will be available for driving the current through the external resistances. But, by Ohm's law, this potential difference is equal to IR.

$$\therefore \quad IR=E-Ir$$

$$\text{or} \qquad I=\frac{E}{R+r}$$

$$\text{or} \qquad \textbf{Current}=\frac{\textbf{E. M. F.}}{\textbf{External resistance}+\textbf{Internal resistance}}$$

Fig. 324 shows a simple circuit in which three resistances of values 6, 2 and 8 ohms (ω is the symbol for ohm) are connected in series. There are four cells, also in series, each of e.m.f.

2 volts, and of negligible internal resistance. Voltmeters are shown connected across the battery and the resistances. An ammeter is included in the cricuit. (Ammeters have negligible resistances).

The total resistance in the circuit is 6+2+8=16 ohms, and the current is 8/16=½ ampere. The drops of potential across the resistances are 3, 1, and 4 volts respectively. The voltmeter connected to the battery will show 8 volts.

Will there be any change in the value of the current if the external resistances change their positions, but remain in series?

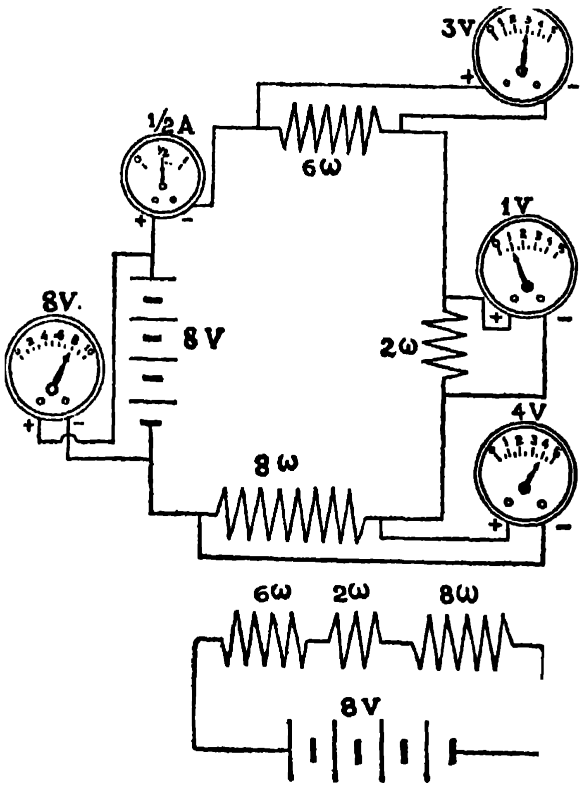

Fig. 324

294. Metre Bridge—Measurement of Resistance

Metre bridge. The value of a resistance can be measured with a high accuracy by means of a metre bridge. In its simplest form (Fig. 325) a metre bridge has three thick strips of copper (the shaded portions *A*, *B* and *C* in the figure) of negligible resistance mounted on an insulating wooden base. They form two gaps, one for the insertion of the resistance to be measured, and the other for another resistance of known value with which it is compared. A straight *uniform* wire, 1 metre long (whence the name

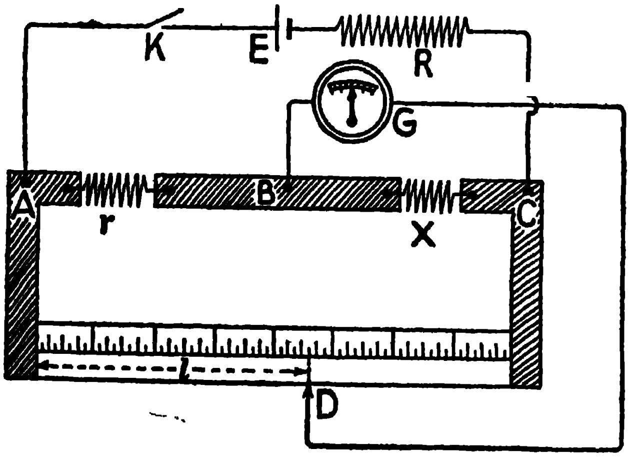

Fig. 325

of the instrument), is stretched between the ends of the strips *A* and *C*. By the side of the wire is fixed a metre scale. A slider (often called a *jockey*) moves through a groove parallel to the wire. With its help contact can be made at any point of the wire. The exact position of contact can be read off from the scale.

Experiment 58. To measure resistance by metre bridge.

Apparatus: Metre bridge, Leclanche or storage cell (*E*), rheostat (*R*), galvanometer (*G*), resistance box, key (*K*), connecting wires (Fig. 325).

Method: Connect as shown in fig. 325. r and X are respectively the known and unknown resistances. Start by making r equal to to 10 ohms. Close the key and slide the jockey until a point of contact *D* is found at which the galvanometer shows no deflection. *D* is called the *null-point*. By adjusting r, bring *D* as close as possible to the middle of the wire. Alter values of r and find the null-point corresponding to each, the null-points preferably lying within the range 40 to 60 cm. Record data as follows.

Serial number	*Resistances in the gaps*		*Position of null-point**	$X=$
	Left	*Right*		
1	r_1	X	l_1	$r_1\times(100-l_1)/l_1$
2	r_2	X	l_2	$r_2\times(100-l_2)/l_2$
3	X	r_3	l_3	$r_3\times l_3/(100-l_3)$
4	X	r_4	l_4	$r_4\times l_4/(100-l_4)$

* [The zero of the metre scale is assumed to coincide with the left end of the wire.]

The mean of the values in the last column may be taken as the value of X. If r is in ohms, X is also in ohms.

Another set of similar readings may be taken by interchanging the connexions at the poles of the cell.

THEORY OF THE METHOD. When there is no current through the galvanometer points *B* and *D* (Fig. 325) are at the same potential. Let I_1 and I_2 be the currents along *ABC* and *ADC* respectively, ρ the resistance per cm of the bridge wire, and l the length of wire to the left of *D*.

$\therefore$ V_{AB} or p.d. between *A* and $B=V_{AD}$ or p.d. between *A* and *D*.

Similarly, V_{BC} or p.d. between *B* and $C=V_{DC}$ or p.d. between *D* and *C*.

Since the resistances of the strips are negligible, we have

$$V_{AB}=I_1 r,\ V_{AD}=I_2 l\rho$$
$$V_{BC}=I_1 X,\ V_{DC}=I_2(100-l)\rho.$$

Hence $I_1 r=I_2 l\rho$

and $I_1 X=I_2(100-l)\rho$

or $r/X=l/(100-l)$

whence $X=r(100-l)/l$

If X is in the left gap and r in the right, we can show in the same way that $X=rl/(100-l)$.

Exercises

1. State Ohm's law in words. Also put it in symbols and explain clearly the meanings of the symbols.

What are the units in which the quantities involved in Ohm's law are expressed ?

2. Define the 'true' ampere, ohm and volt, and also the 'international' ampere, ohm and volt. What is the advantage of the latter definitions ?

3. How does the resistance of a conductor depend on its length, diameter and material ? Of two wires of the same material one is double as long as the other, and also thrice as thick. What is the ratio of their resistances ? (*Ans:* 2:9.)

4. What is resistivity ? How is it expressed ? How does it depend on temperature ?

5. State how resistances change when they are joined (*i*) in series, (*ii*) in parallel.

You are given three resistances of values 1, 2 and 3 ohms. Any two of them may be connected either in series or in parallel. Say which pair will have the highest resistance, and which the lowest. State also the nature of the connexion in the pair.

6. What is a shunt ? A current of 5 amperes flows through a resistance of 5 ohms. It is proposed to reduce this current to 1 ampere by using a shunt. What should be the resistance of the shunt ? (*Ans* : 1·25 ohms.)

7. A battery consists of three Leclanche cells each of e.m.f. 1·5 volts and resistance 3 ohms. Find the e.m.f. and internal resistance of the battery when the cells are joined (*a*) in series, (*b*) in parallel.

[*Ans* : (*a*) 4·5 volts, 9 ohms ; (*b*) 1·5 volts, 1 ohm.]

8. A voltmeter connected to a battery indicates its e.m.f. as 10 volts. But when the battery drives a current of 2 amperes through an external resistance, the voltmeter indicates 9 volts. Explain the difference. What is the internal resistance of the battery ? What is the resistance of the external circuit ?

(*Ans* : $\frac{1}{2}$ ohm ; $4\frac{1}{2}$ ohm).

9. Two resistances, 5 ohms and 10 ohms, are in series. A potential difference of 150 volts is applied across the two. Find the current and also the p.d. across each.

(*Ans* : 10 amp ; 50 and 100 volts).

10. Two resistances, 10 and 20 ohms respectively, are in parallel. They carry a total current of 6 amperes. What are the current and the p.d. across each ?

(*Ans* : 4 amperes and 40 volts in the case of the lower resistance ; 2 amperes and 40 volts for the other.)

11. A potential difference of 80 volts is applied to a circuit consisting of a wire of 7 ohms in series with a parallel arrangement of a 4-ohm and a 12-ohm wire Find the current and p.d. across each wire.

(*Ans* : 7-ohm wire—current=8 amp, p.d.=56 volts ; 4-ohm wire—6 amp and 24 volts ; 12-ohm wire—2 amp and 24 volts.)

12. Three wires of 1, 4 and 5 ohms in parallel carry a total current of 14·5 amperes. What is the current in each ? What is the p.d. across the common terminals ?

(*Ans* : 10, $2\frac{1}{2}$ and 2 amp respectively ; 10 volts.)

MAGNETIC EFFECTS OF CURRENTS

295. Magnetic Field due to a Straight Current

In Experiment 53 (§274) we have seen that a conductor carrying a current produces a magnetic field around it. To find the nature of the magnetic field surrounding a straight current (i.e. a current in a straight conductor), bore a hole in a piece of cardboard or glass, fix it in a horizontal position, pass a long, straight wire vertically through it and send a current of about 10 amperes through the wire. Sprinkle iron filings on the board and tap gently. The filings will arrange themselves in concentric circles round the wire and close to it (Fig. 326a). These indicate the lines of force in the magnetic field due to the current.

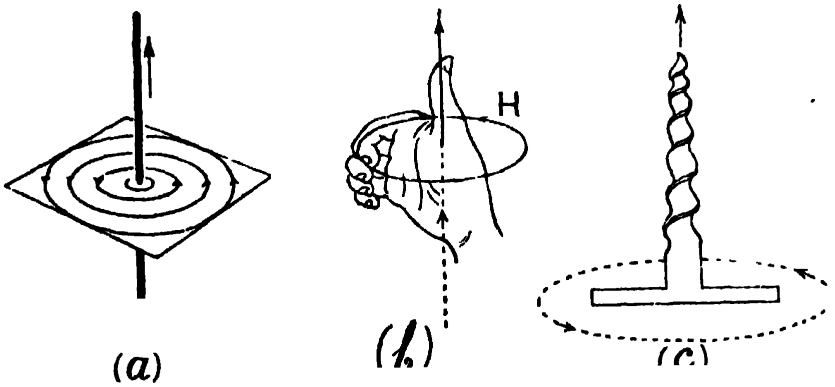

Fig. 326

Place a small magnetic needle close to the wire and note the direction in which the N-pole points. This gives the direction of the lines of force of the field created by the current. An easy way of finding this direction will be to apply what is known as the **Right hand thumb rule** (Fig. 326b). Close the fingers of the right hand around the outstrectched (right) thumb. If the thumb points to the direction of current flow the direction in which the fingers point will give the direction of the field. The same relation is also given by **Maxwell's cork-screw rule** (Fig. 326 c), which may be stated as follows :

If an observer imagines that a cork-screw is being driven in the direction of the current, a north pole, placed in the field, will move in the same direction as the screw is being turned.

The behaviour of the magnetic needle in Experiment 53 may be understood on the basis of any of these rules. The N-pole deflects in the direction of the field. From fig. 327 it is easy to find how the N-pole will be deflected in the cases represented there.

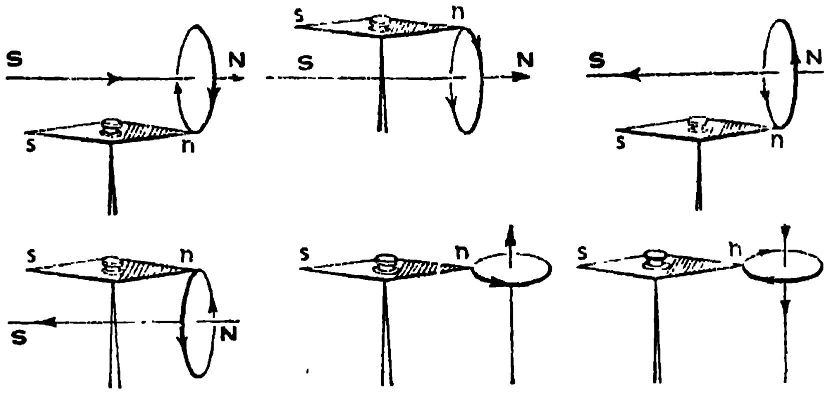

Fig. 327

Note that lines of force round a straight current are closed circles concentric with the current. They have no beginning or ending. No free polarity appears anywhere. The conductor itself is not magnetized; it does not attract iron filings. The surrounding region becomes a magnetic field.

296. Magnetic Field due to a Circular Current

Pass two ends of a stout copper wire through two holes, a few inches apart, in a piece of glass or cardboard. Bend the wire into a circular form (Fig. 328) and pass a strong current (about 10 amp.) through it. Sprinkle iron filings on the board and tap gently. The filings align themselves in the direction of the resultant magnetic field due to the coil and the earth. Near the wires the lines of force are circular. If the plane of the coil is perpendicular to the magnetic meridian, the lines at the centre will be normal to the plane of the circle.

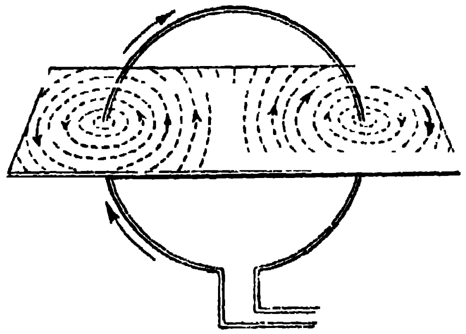

Fig. 328

The direction of the field at the centre of the coil or at any other place may be obtained from the right hand thumb rule.

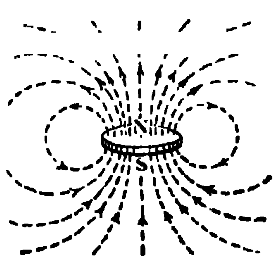

Fig. 329

297. Circular Current and Circular Magnetic Shell

A sheet of magnetic material magnetized perpendicular to its faces is called a magnetic shell. Fig. 329 shows a circular magnetic shell and its lines of force. The shell has north polarity on one face and south polarity on the other. A comparison of the lines of force due to this shell and to the circular coil will show that they are similar. A circular coil carrying a current behaves like a magnetic shell. The two faces of the coil have opposite polarities. The following rules gives the polarity on a face :

If, looking at one face of the coil, the current appears to flow anticlockwise, the face will have north polarity ; if clockwise, the face will have south polarity (Fig. 330).

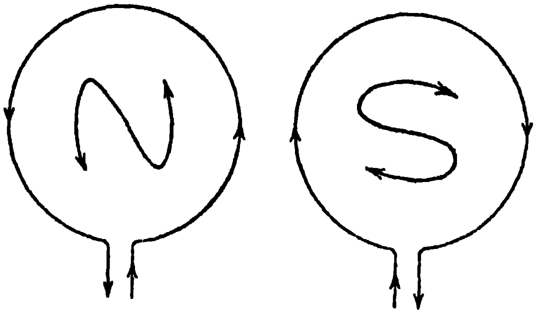

Fig. 330

De la Rive's experiment. An interesting demonstration of the magnetic behaviour of a coil carrying a current is due to de la Rive. The ends of a small coil of several turns of insulated copper wire are soldered to a strip of copper and zinc respectively. The strips pass through a large cork which is floated in dilute sulphuric acid (Fig. 331). A current passes through the coil from copper to zinc.

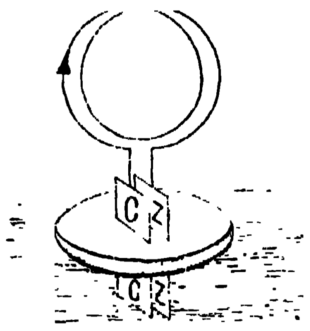

Fig. 331

If now the N-pole of a magnet is presented to the face of the coil in which the current appears to flow anticlockwise, the coil is repelled. When the south pole is so presented, there is attraction.

Left to itself the coil sets with its plane perpendicular to the magnetic meridian. On looking at its north face the current appears to flow anticlockwise.

Each turn of the coil behaves like a magnetic shell. The effect of having several turns is to increase the effect. The entire coil behaves as a bar magnet with a short axis in the direction of the axis of the coil.

298. Magnetic Field due to a Solenoid

A coil of many turns of wire wound closely on a cylinder is called a solenoid. (The flow of current is everywhere perpendicular to the axis of the cylinder.)

To investigate the magnetic field due to a solenoid, cut three sides of a rectangle in the middle of a cardboard and pass the free end of the strip, so secured, through the solenoid (Fig. 332). Sprinkle iron filings on the board, pass a strong current through the solenoid and tap the board gently. It will be observed that

(*i*) the lines of force outside the solenoid are similar to those of a bar magnet, and

(*ii*) inside the solenoid the lines are crowded together in a direction parallel to the axis.

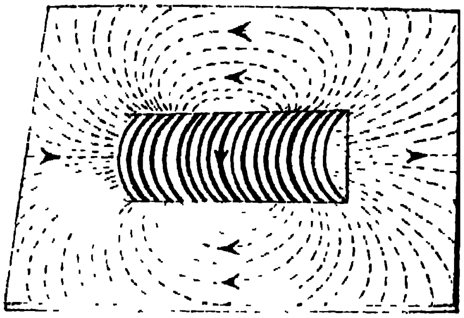

Fig. 332

Fig. 333 shows the lines due to a solenoid and a bar magnet side by side. To show clearly the field inside the solenoid, its turns have been shown much separated in the diagram.

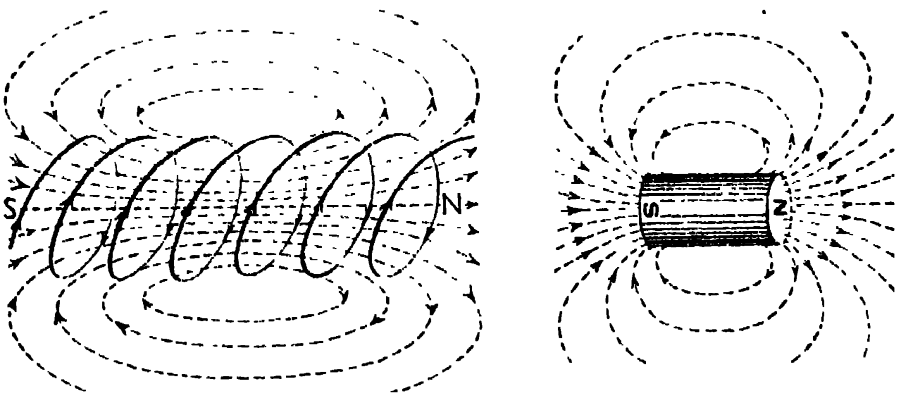

Fig. 333

It is clear that a solenoid is equivalent to a number of coils arranged coaxially and behaves like a bar magnet. When floated on cork as in de la Rive's experiment, it shows a more pronounced effect. The polarity is given by the same law as for a circular current.

299. Electromagnets

The discovery of electromagnets is credited to William Sturgeon who, five years after Oersted's discovery, filled the interior of a solenoid with soft iron and produced a powerful magnet.

Whenever a bar of soft iron is inserted in a solenoid through which a current is passing, the bar becomes strongly magnetized by the field within the solenoid. The combination of the soft iron bar and the solenoid is called an electromagnet. The polarity of the electromagnet is determined by that of the solenoid. If, on looking at one end of the solenoid the current in it appears to flow anticlockwise, this end of the electromagnet will have north polarity. Also see § 226.

The strength of an electromagnet depends on the current through the solenoid and the total number of turns in it. This product is called *ampere-turns.* The strength cannot, however, be increased beyond a certain value depending on the quality of iron.

300. Electric Bell

The electric bell provides a simple application of the electromagnet. Its construction and mode of action are described below.

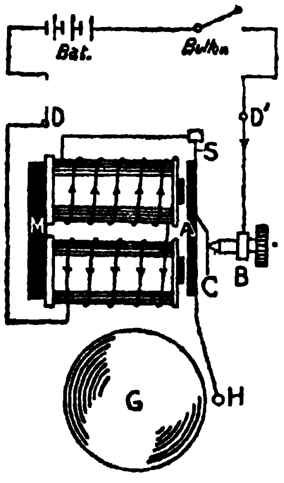

Fig. 334

In fig. 334, *M* is an electromagnet which is excited by a current from a battery. The current flows through the coil surrounding *M* when the button of the switch in the battery circuit is pressed. *CS* is a spring which is normally in metallic contact with an adjustable screw *B*. Attached to *CS* there is a piece of soft iron *A* which is attracted towards *M* when the current passes. As *A* is drawn towards *M* the contact between *CS* and *B* breaks and the current stops. As the magnet is no longer excited, *CS* flies back to its position and starts the current again. *M* is again excited, draws *A* away and breaks the circuit. The process is repeated so long as the button of the switch remains pressed. Attached to *CS* there is a hammer *H* which strikes the gong *G* and continues to do so until the button is released.

Sparking occurs between *B* and its point of contact with *CS* when the circuit breaks there. To minimize its effect *B* may be provided with a platinum point and *CS* with a platinum stud at the point of contact.

301. The Telegraph

The electromagnet found an early and important application in the telegraph. In its simplest form a telegraph equipment consists of :

(*i*) A **circuit,** which is a line of wire connecting two stations. The earth provides the return path.

(*ii*) A **battery** to supply the current for flow along the line. The current is interrupted, the interruptions making up the code of the message.

(*iii*) A **key** or **transmitter** for opening and closing the circuit at the transmitting end.

(*iv*) A **sounder** or **receiver** at the receiving end. It responds to the interrupted current.

(*v*) **Relays**, for work on long telegraph lines.

The **key** or **transmitter** is a simple contact breaker so designed that it can be moved several hundred times a minute. The way it is connected with the battery and the circuit is shown in fig. 335. When the knob *K* of the key is depressed the terminal *P* of the battery is connected with the line *L*. A current flows through the line, the return path is provided by the earth to which one pole of the battery is connected.

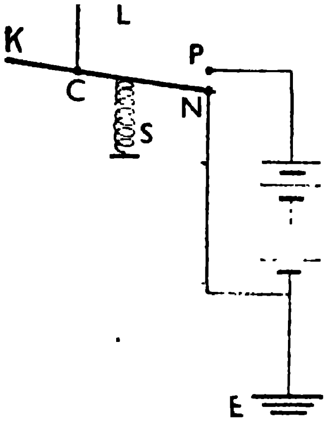

Fig. 335

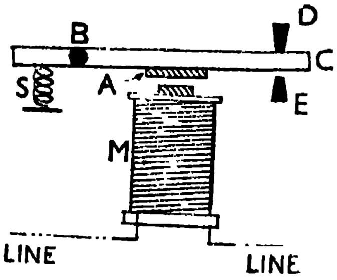

Fig. 336

The **sounder** or **receiver** has an electromagnet *M* (Fig. 336). When the line-current passes through its coils, it attracts a soft iron armature *A* and brings the end *C* of a lever sharply against a stop *E*. When there is no line current the spring *S* drives *C* against the stop *D*.

The sounder responds to the interrupted current coming from the transmission end. When the key of the transmitter is depressed, the armature of the receiver is also depressed. When the former is released the latter is released too.

When a short duration current passes through *M*, *C* strikes against *E* and *D* in quick succession and produces two sharp clicks separated by a small interval. For a current of longer duration, the interval between the clicks is longer. The short signals are called *dots* and the longer ones, *dashes*. The letters of the alphabet are made up by a combination of dots and dashes according to an accepted code. The one devised in 1837 by the American inventor, Samuel Morse, is widely used.

Fig. 337 shows a simple telegraph circuit with a transmitter and receiver. It can send a message in one direction only.

The Relay. In long telegraph lines the current is so weakened by the resistance of the line and by leakage that it becomes too weak to operate the sounder. In such a case the weak current is made to bring into operation a local current strong enough to work the sounder. A device used for such a purpose is called a relay.

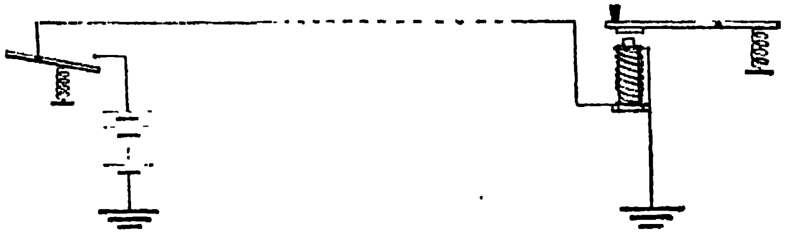

Fig. 337

The electromagnetic relay is very nearly the same in construction as a sounder. Its electromagnet *M* (Fig. 338) is strong enough to pull the armature *A* even when the current is feeble. When the lever strikes the stop *E* a local battery circuit is closed and the electromagnet in the sounder is excited.

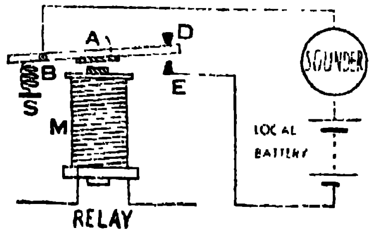

Fig. 338

302. Moving Magnet Galvanometers

A *galvanometer* is an instrument for detecting and measuring small currents. When it is modified so as directly to read the value of a current it is called an *ammeter.*

In one type of galvanometer there is a fixed coil consisting of a number of turns of insulated copper wire through which the current to be detected passes (Fig. 339). The plane of the coil is in the magnetic meridian. A small magnetized needle is pivoted or suspended at the centre of the coil. The needle is acted upon by two magnetic fields—one due to the coil which deflects the needle and tends to set it at right angles to the coil, and the other due to the earth's field which tends to restore it to its position. Obviously, the larger the current the larger the deflection.

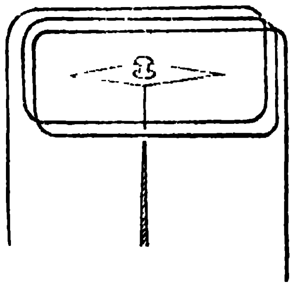

Fig. 339

This type of galvanometer is known as the *moving magnet* type. A *tangent galvanometer* is one of this type. Its coils are circular.

303. Mechanical Force on a Current in a Magnetic Field

Faraday's experiment. Faraday was the first to show that a mechanical force acted on a wire carrying a current when the wire was placed in a magnetic field. In his experiment, a conductor was suspended at its upper end from a hook and dipped at its lower end into a trough of mercury (Fig. 340). One pole of a strong bar magnet projected a little through the mercury. When a current was passed through the conductor it rotated continuously around the magnet. If the current moved downwards, then looking in the direction of the current from above, the conductor would appear to move clockwise around a north pole. On reversing the current or the polarity of the magnet, the sense of rotation was also reversed. Faraday's experiment was the forerunner of the modern electric motor.

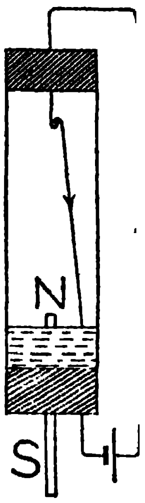

Fig. 340

Fleming's left-hand rule. The direction of the mechanical force on the conductor is given by Fleming's left-hand rule It may be follows :

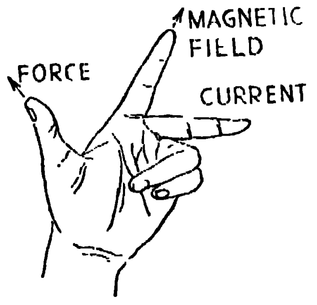

Fig. 341

*Stretch the thumb, forefinger and middle finger of the left-hand at right angles.** *If the Forefinger points to the direction of the magnetic Field and the middle finger to the direction of the current* (I), *the thuMb will give the direction of the force on the conductor* (*hence of the Motion*) (Fig. 341).

Magnitude of the force. If the current flows perpendicular to the magnetic field, the force on a conductor of length L cm will be $HIL/10$ dynes where H is the strength of the field in oersteds and I, the current through it *in amperes*. The force is zero when the current and the field are parallel. In general, when the angle between them is α, the force is $HIL \sin \alpha/10$.

Barlow's wheel. A common, laboratory experiment to demonstrate the mechanical force on a current in a magnetic field is known as Barlow's wheel experiment. A small copper wheel, capable of easy rotation about an axis through its centre, is mounted between the poles of a strong horseshoe magnet (Fig. 342). Its rim dips into

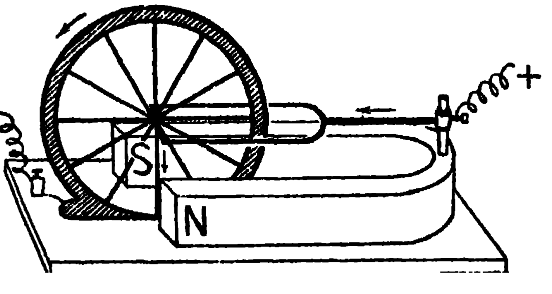

Fig. 342

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* Try supporting a dish on these three fingers

a small trough of mercury, which is connected to one pole of a battery. The other pole is connected to the centre of the wheel through its axis. The current, therefore, flows radially along the spokes.

When the current flows the wheel rotates continuously. The direction of the force on a spoke which carries the current from the axis to the rim can be found with the help of Fleming's left-hand rule. In the set up represented in fig. 342 the wheel will be found to rotate in an anticlockwise directon when viewed from the side of the N-pole. Reversing the current or the poles of the magnet will reverse the sense of rotation.

To obtain the best effect the lines of force of the magnetic field should be perpendicular to the plane of the copper wheel.

The wheel may be a solid disc. Sometimes the disc is made in the shape of a star.

304. The Electric Motor

Principle. An electric motor is a machine which converts electrical energy into mechanical energy. Its principle is illustrated in fig. 343. A loop *ABCD* of wire carrying a current is placed in a magnetic field as shown. An application of Fleming's left-hand rule shows that *AB* will be urged upwards and *CD* downwards by the mechanical forces acting on them due to the magnetic field. If the loop is free to turn about an axis through its centre parallel to *AB*, it will rotate until it is just midway between the poles. While in this position the current through it is reversed in direction by means of sliding contacts and a split ring commutator. This reverses the direction of the forces on *AB* and *CD*. *AB* is now urged downwards and *CD* upwards. The loop therefore rotates through another half-turn. The sliding contact and commutator arrangement reverses the current through the coil at every half turn. This keeps the loop rotating continuously.

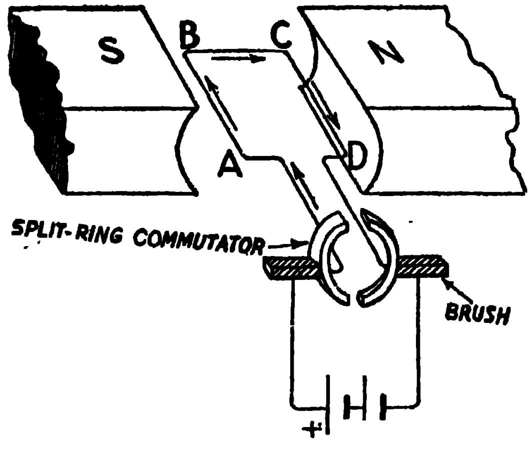

Fig. 343

Construction. The construction of a simple commercial motor is illustrated in fig. 344. Several separate loops of wire are mounted at different angles on a drum, the ends of each loop being connected to two small copper bars insulated from each other and arranged in a ring called the *commutator*. The drum (with the coils), and the commutator are mounted on a shaft which rotates on smooth bearing in the field of an electromagent. The rotating part is called the *armature*. The current from a source is fed into the coils through 'brushes'. The arrangement is such that all conductors opposite one pole are urged upwards, and those against the other pole are urged downwards. The source that supplies current to the coils also excites the electromagnet by sending a current through the field coils. The drum is made of laminated sheets of iron insulated from each other.

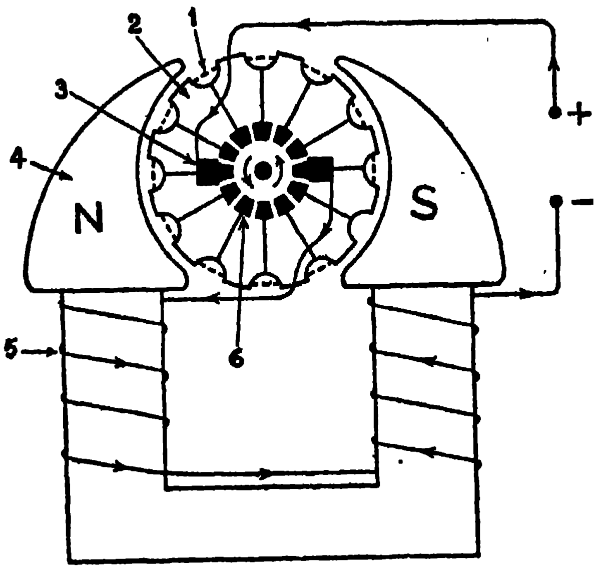

Fig. 344
1—Conductors; 2—Drum armature; 3—Brushes; 4—Poles of electro-magnet; 5—Field coils; 6—Commutator bars.

305. Moving Coil Instruments—Ammeters and Voltmeters

We have seen in connexion with the electric motor that a loop of wire carrying a current tends to rotate when placed in a magnetic field. Many electrical measuring instruments, such as ammeters, voltmeters and galvanometers, depend for their action on this effect. They are called *moving coil* instruments. The instruments are, however, much more delicate in construction than the motor, and the motion of the loop of wire is restricted by springs, as in commercial ammeters and voltmeters, or by the twist of the wire which supports the loop as in galvanometers.

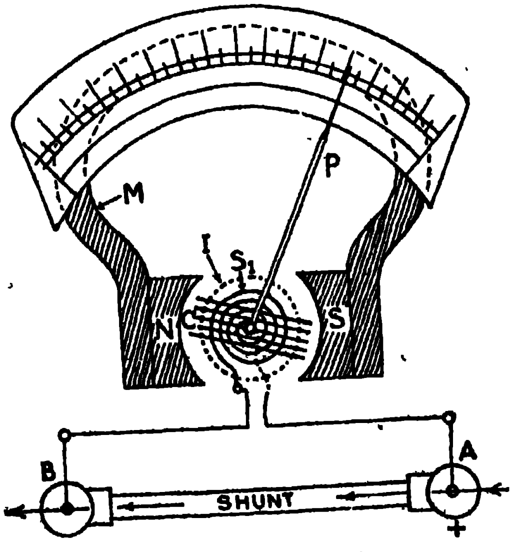

Fig. 345

Ammeters and Voltmeters. In commercial ammeters and voltmeters (Figs. 345 and 346), a coil C of fine wire is mounted between the two poles N, S of a permanent magnet M and its motion is restrained by one or more delicate springs (S_1). The further the coil moves from its position of rest the greater the restoring force exerted on it by the springs. A long pointer P attached to the coil moves over a scale graduated in amperes or volts as the case may be. The larger the current through the coil the greater its rotation.

An ammeter is connected in series with the conductor the current through which it wants to measure (Fig. 347). Ammeters must have a low resistance so that its insertion in a circuit may not materially alter the current. Most ammeters have a low resistance, called a *shunt*, connected across its coil as shown in fig. 345. By altering the value of the shunt the same instrument can be used over different ranges.

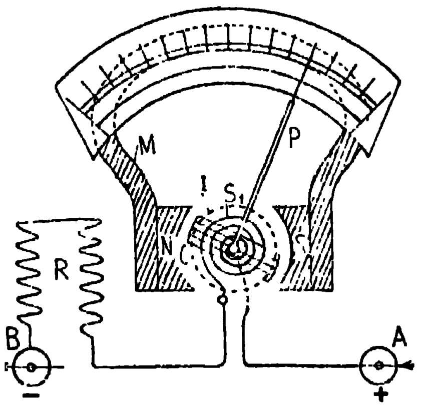

Fig. 346

A voltmeter is connected in parallel with the conductor the difference in potential between the terminals of which it wants to measure (Fig. 347). A voltmeter must have a high resistance. It must not draw a large current from the circuit. Unless the current through the voltmeter is small, its introduction in the circuit will materially alter the potential difference intended to be measured. To give a voltmeter a high resistance, a series resistance R (Fig. 346) is connected with the coil. By altering the value of R the same instrument may be used over different ranges.

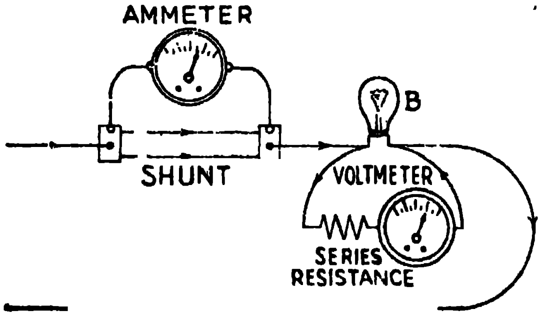

Fig. 347

In both instruments the positive (+) terminal must be connected to the positive (+) side of the circuit.

306. Effect of Current on Current

A conductor carrying a current produces a magnetic field around it. Also, a mechanical force acts on a conductor carrying a current when the latter is placed in a magnetic field. Hence, if two conductors, each carrying a current, are placed near each other, one will exert a force on the other.

The following results can be derived from experiment, or from Fleming's left-hand rule:

(*i*) Currents flowing in the same direction in parallel wires attract each other.

(*ii*) Currents flowing in opposite directions in parallel wires repel each other.

(*iii*) Currents moving towards or away from a point attract each other.

(*iv*) When one current moves towards a point and the other away from it, a repulsion occurs between the two.

The results are represented in fig. 348.

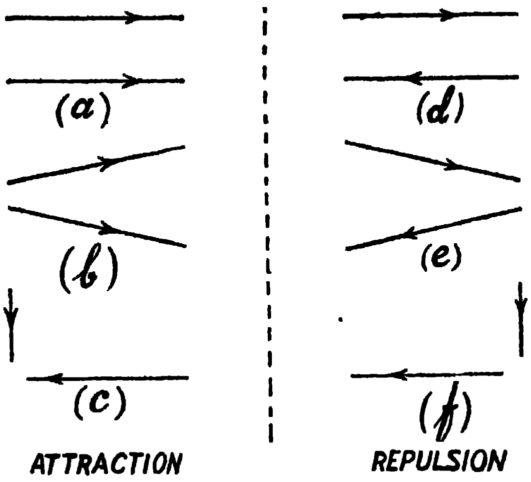

Fig. 348

Roget's vibrating spiral. A striking demonstration of the attraction between parallel currents is provided by Roget's vibrating spiral (Fig. 349). It consists of a long coil of copper wire, a few centimetres in diameter. The upper end is fixed and connected to a binding screw. The lower end, which terminates in a metal ball, just touches the mercury in a cup. When a current is passed through the coil, it flows in parallel directions through successive turns. Each turn, therefore, attracts the other. The coil shortens as a result, and the circuit breaks. With it the attraction ceases. The coil regains its length and establishes the current on touching the mercury. Contraction and expansion of the coil succeed each other and the spiral goes on vibrating so long as the current passes.

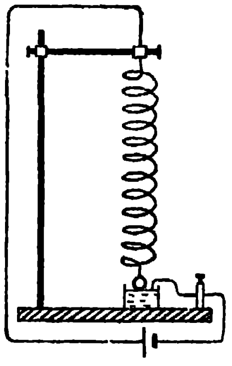

Fig. 349

Each turn of the coil may also be looked upon as a magnetic shell with opposite polarities on two sides. The neighbouring faces of two adjacent shells will have opposite polarities. Each will, therfore, attract the other.

Exercises

1. How would you demonstrate to a class that an electric current produces a magnetic field around it?

State the laws connecting the directions of the magnetic field and the current.

Does the conductor carrying the current behave as a magnet and attract iron filings?

2. Draw diagrams showing the similarity between the magnetic fields produced by a circular current and a magnetic shell. State the law determining the polarity of a circular current.

Describe a simple experiment to show that a solenoidal current is equivalent to a bar magnet.

3. What is an electromagnet? What is meant by *ampere-turns*? Can the strength of an electromagnet be increased as much as we like? Should the coils round the two limbs of a U-shaped electromagnet be wound in the same way? Why?

4. Draw a neat sketch of an electric bell and label the parts. Explain how the make-and-break arrangement acts.

5. Briefly describe the action of a telegraphic transmitter and receiver.

What is transmitted along the line?

What is a relay? How does it act?

6. How can you show that a mechanical force acts on a current in a magnetic field? State Fleming's left-hand rule in this connexion.

What will be the force if the current and magnetic field are parallel?

7. Draw a simple diagram to illustrate the action of an electric motor. What part does the commutator play in its action?

8. For what purposes are ammeters and voltmeters used? Show by means of a diagram how they are connected to a circuit.

Why should the ammeter have a low resistance and the voltmeter, a high resistance? If both have moving coils of the same kind, how are the two made to have very different resistances?

9. What kinds of action can take place between currents in parallel wires? Explain the action of Roget's vibrating spiral.

CHAPTER 36

HEATING EFFECT OF CURRENTS

307. Joule's Law of Heating

The flow of an electric current through a conductor always produces heat. The law relating to the heating effect of currents was experimentally investigated by Joule, who found that **the heat generated by a current flowing through a conductor is proportional to (i) the square of the current, (ii) the resistance of the conductor and (iii) the time for which the current flows.** This statement is known as **Joule's law of heating.**

Experiment shows that the passage of a current of I amperes through a resistance of R ohms for time t seconds will generate H calories of heat, where

$$H = \cdot 24\, I^2 Rt.$$

If V is the potential difference in volts between the terminals of the resistance, $V = IR$. We may, therefore, write

$$H = \cdot 24\, I^2 Rt = \cdot 24\, VIt = \cdot 24\, V^2 t/R \qquad (307 \cdot 1)$$

308. Experimental Verification of Joule's Law

Joule's law can be verified by immersing a coil of known resistance in a non-conducting liquid of known specific heat contained in a calorimeter, and sending a current through it. Fig. 350 shows the experimental arrangement. B is a battery which drives a current through the coil R of known resistance immersed in water in a calorimeter. The rheostat r controls the value of the current. This value can be read off from the ammeter A connected in series with R. V is a voltmeter connected in parallel with R. It gives the potential difference between the ends of R. The thermometer T gives the temperature of the liquid. K is a key which starts or stops the current as necessary.

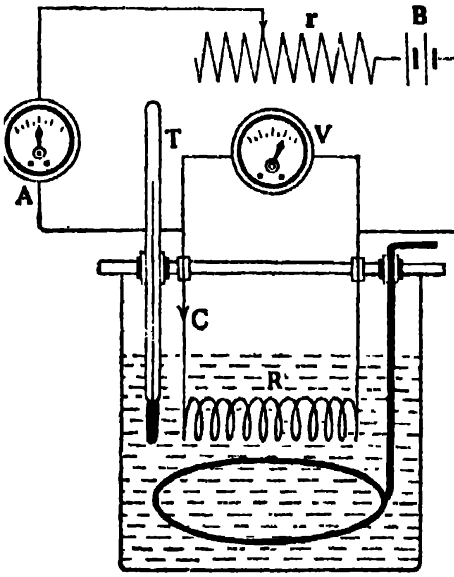

Fig. 350

(i) **To verify the law of currents.** To show that the heat developed in R by a current I flowing through it is proportional to I^2 when R and t are constant, we proceed as follows :

Note the initial temperature of water. Send a current through it and stir the liquid briskly. Stop the current, say, after five minutes. Note the final temperature. During this period the current should be kept constant by adjusting the rheostat, if necessary. The value of the current is noted from the ammeter, and the time, from a stop-watch.

If W is the water-equivalent of the calorimeter and its contents, and Θ the temperature rise, then the heat developed is $W\Theta$.

Allow the calorimeter and its contents to cool to the room temperature. Increase the current by about 10% and repeat, taking care that the current flows for exactly the same time as before. Note the rise of temperature. If the two cases are distinguished by subscripts 1 and 2, it will be found that

$$\frac{H_1}{H_2}=\frac{W\Theta_1}{W\Theta_2}=\frac{\Theta_1}{\Theta_2}=\frac{I_1^2}{I_2^2}$$

(ii) **To verify the law of time.** Keep the current the same, but alter the time of flow. It will be seen that

$$\frac{H_1}{H_2}=\frac{\Theta_1}{\Theta_2}=\frac{t_1}{t_2}$$

(iii) **To verify the law of resistance.** In the two sets of observations use two coils of different resistances, say, R_1 and R_2. Keep I and t the same in both experiments. It will be found that

$$\frac{H_1}{H_2}=\frac{\Theta_1}{\Theta_2}=\frac{R_1}{R_2}$$

The ratio of the voltmeter reading to the ammeter reading may be taken to give the value of R.

309. Determination of the Mechanical Equivalent of Heat

Electricity, in moving from a place of higher potential to a place of lower potential, does work. When a current of I amperes flows through a conductor of resistance R ohms for t seconds, the charge transported is $I\times t=Q$ coulombs. Between the ends of the conductor the potential difference is $I\times R=V$ volts. The total work done is, therefore, VQ joules (See § 267).

This work appears as heat in the conductor. Let H calories of heat be generated in this way. The mechanical equivalent of heat, J, is the amount of work that must be done to produce 1 calorie of heat. From the above we shall have

$$HJ = VQ$$

or
$$J = \frac{VQ}{H} = \frac{VIt}{H} = \frac{I^2Rt}{H}\frac{(\text{joules})}{(\text{calories})}$$

Comparing this relation with Eq. 307.1 we find that the constant ·24 in that equation is the reciprocal of J. J has the value 4·2 joules per calorie. (Also see § 139). With the arrangement described in § 308 it is easy to find the value of J after measuring I, R, t and H. Electrical method of measuring J is one of the most accurate.

Example. In an actual experiment for the determination of J, the following values were observed : $V = 76{\cdot}0$ volts, $I = 1.20$ amp, $t = 525$ sec, water-equivalent of calorimeter and contents $= 1080$ gm, rise of temperature $= 10{\cdot}5°C$. To find J.

Solution. The work done $= VIt = 76 \times 1{\cdot}2 \times 525 = 4{\cdot}788 \times 10^4$ joules.

Heat developed $= 1080 \times 10{\cdot}5 = 1{\cdot}134 \times 10^4$ calories.

$$J = \frac{\text{Work}}{\text{Heat}} = \frac{4{\cdot}788 \times 10^4 \text{ joules}}{1{\cdot}134 \times 10^4 \text{ calories}} = 4{\cdot}22 \text{ joules per calorie.}$$

310. Units of Electrical Power and Electrical Energy

Unit of electrical power. Power has been defined as work done per unit time (§ 70). If I is the current in amperes flowing through any portion of an electrical circuit, and V is the voltage (potential difference in volts) across the portion of the circuit, the power or the rate at which energy is being dissipated in the portion of the circuit is equal to VI watts.

The **watt** is the power or the energy expended per second when a current of one ampere flows under a potential difference of one volt.

Watts = Volts × Amperes

1 kilowatt (kw) = 1,000 watts.

Unit of electrical energy. Since power × time = work, an agent delivering a power P continuously for time t does a total amount of work equal to the product of P and t. This leads to the definition of various units of electrical work or energy.

When the power is one watt, the energy delivered or consumed in one second is one **joule.**

Joules = Watts × Seconds.

The energy delivered or consumed in one hour when the power is one watt, is called a **Watt-hour.**

$$\text{Watt-hours} = \text{Watts} \times \text{hours}$$

$$1 \text{ watt-hour} = 1 \text{ watt} \times 1 \text{ hour} = 1 \text{ watt} \times 3600 \text{ sec} = 3600 \text{ joules.}$$

A still larger unit of electrical energy has been chosen as the *legal unit* by the Board of Trade for commercial purposes. It is called the **kilowatt-hour** (kwh) or the **Board of Trade Unit** (B.O.T. unit) and is equal to 1000 watt-hours. It is thus the *electrical energy supplied to an electrical circuit in one hour when the power in it is one kilowatt.*

$$\text{Kilowatt-hours} = \frac{\text{Watt-hours}}{1{,}000} = \frac{\text{volts} \times \text{amperes} \times \text{hours}}{1{,}000}$$

When power is expended in overcoming a resistance R, its value is RI^2. It degrades into heat and is known as RI^2 loss.

Example. A street cable has a resistance of ·05 ohm and carries a current of 100 amperes. Find in kilowatt-hours the energy lost in it as heat in 24 hours.

Solution: Number of kilowatt-hours $= \frac{\text{watt-hours}}{1{,}000}$

$$= \frac{\text{volts} \times \text{amperes} \times \text{hours}}{1{,}000} = \frac{(\cdot 05 \times 100) \times 100 \times 24}{1{,}000} = 12$$

311. Some Applications of the Heating Effect of Currents

Applications of Joule heating are too many to enumerate. We shall consider only a few.

A. Electric heaters. Such diverse household appliances as the electric kettle, electric iron, toaster, coffee percolator, electric oven, radiator etc. may be included under this general head. All these heating devices contain heating coils, usually made of wires or strips of *nichrome* (an alloy of 60% nickel, 24% iron and 16% chromium) which has a high specific resistance and can be raised to a red heat and maintained there for a long time without damage to the wire. The coils are wound on mica or fire-clay, which is not only electrically insulating but can also withstand heat.

Each appliance is meant for use at a fixed voltage, which must be adhered to. This voltage and the power of the appliance at this voltage are imprinted on it. The stamp '220-volt, 500-watt' on a heater means that it should be used on 220 volt lines, and when so used it will consume power at the rate of 500 watts. The current it will draw at the marked voltage will be 500/220=2·27 amperes nearly. Its resistance is 220/2·27 ohms.

B. Fuses. The current in a circuit may rise to enormous values when its opposite terminals are brought together or are connected by a low resistance through carelessness or accident. When this takes place, excessive heating is produced in the circuit causing damage to it or any machine or appliance which it may include. To guard against an emergency of this kind a simple device, called a **fuse**, is connected to the circuit *in series*. It is a short length of wire or a strip of suitable material and dimensions which melts and breaks the circuit if the current through it exceeds a predetermined value. A '5 ampere fuse' is one which melts and breaks the circuit when the current in the fuse exceeds 5 amperes.*

Fuses are always fixed in supports of fire proof material, usually glazed porcelain. Tin, lead, an alloy of tin and lead, strip zinc, copper and aluminium are some of the materials used for fuses. For fuses of a given material, the thinner it is the smaller the current required to melt it.

C. Electric furnaces. Temperatures of 1,000°C or more can be reached in electric furnaces. The current is sent through a nichrome or platinum wire wound round a heat-resistant, non-conducting tube, and lagged on the outside with a thick layer of heat-insulating material. Hardening, annealing and enamelling are often done in such furnaces.

In some industrial electric furnaces the current is sent through the materials to be reacted, e.g., in the manufacture of calcium carbide, carborundum, calcium cyanamide (a fertilizer), extraction of zinc, manufacture of special alloy steels containing tungsten or molybdenum etc.

D. Welding. In *resistance welding*, the two metals to be welded are clamped together and a very strong current passed through them across the surface of contact. The junction, which is practically the seat of the resistance, is raised to white heat by the current. The metals then join under pressure.

In *spot welding* the overlapping metals are clamped into close contact, a metal electrode pressed against each side and a heavy current passed from one electrode to the other. The heat produced causes the plates to join along the line of the electrodes.

E. Electric filament lamps. As the current through a wire increases, the wire gets hotter. At about 550°C it emits a dull red light. The colour passes through bright cherry at 1000°C to bright orange at 1200°C and white at 1300°C. At 1500°C and above the colour is dazzling white. The higher the temperature the more and the brighter the light emitted.

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* It is not practicable to design a fuse which will break a circuit on a very small overload. There is a fairly wide margin between the current a fuse will safely carry and the one that will melt it.

In the modern electric filament lamp there is an exceedingly fine wire of metallic tungsten heated to a temperature of over 2000°C by an electric current. There are two main types, *viz.*, (i) the *vacuum* lamp and (ii) the *gas-filled lamp*. The bulb of the former is evacuated. The running temperature is about 2100°C. This type is becoming obsolete. The gasfilled lamp contains argon with a little nitrogen at a pressure of about three-fourths of an atmosphere at the room temperature. The running temperature is about 2700°C. Cooling of the filament is reduced to negligible proportions by winding the filament into a fine spiral. The length of the spiral is less than one tenth of the total length of the wire. The spiral is suspended from radial arms in a horizontal plane. In the **coiled-coil filament,** the single spiral is itself coiled into a second spiral. It gives about 15 to 20 per cent more light for the same power consumption.

The *luminous efficiency* of a lamp is the ratio of its candle power to the watts it absorbs. For a coiled-coil lamp it is about 2 candles per watt. The lamps are rated in watts. The voltage at which a lamp is designed to operate is also specified on it. A '60-watt, 220-volt' lamp will give about 100 c.p. when operated at 220 volts. At 110-volts it will give no light. A 110-volt lamp connected to 220-volt mains will burn out. The life of a lamp operated at the specified voltage is about 1000 hours.

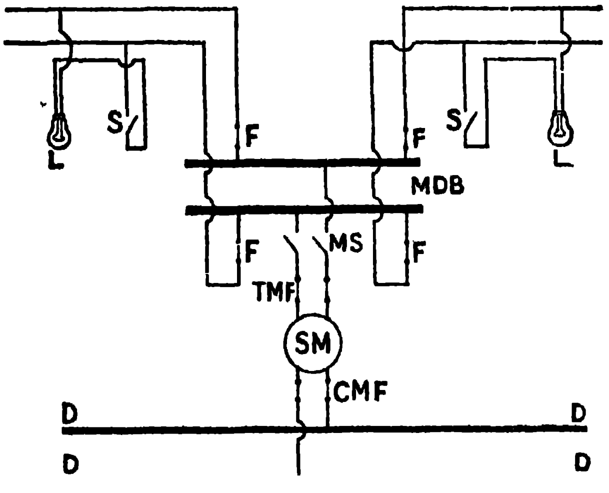


Fig. 351

Emission of light is sensitive to changes of voltage. An increase of 5% in the voltage will increase the emission of light by about 25%. Continued operation at a voltage higher than the one specified, will reduce the life of a lamp.

312. Indoor Wiring of Buildings

It is useful to have an idea of the indoor wiring of a small building. Fig. 351 represents a simple arrangement. *DD* are the street distribution mains from which service mains pass into the house. Fuses *CMF*, known as company's main fuse, are placed on sevice mains. The mains are then led into the supply meter *SM* which records the energy consumed in the household. *TMF* is the tenant's main fuse, and *MS* the main switch. *MDB* is the main distribution board. From it cables run to the different floors, fuses *F* being inserted as shown (one for each cable). The diagram shows only two circuits from the main distribution board. There may be more ; but all are connected in the same way as those shown. The lamps or other electrical appliances, such as fans etc., are connected *in parallel* between the mains. Each lamp (*L*) is controlled by a switch (*S*).

313. Worked Examples

(1) A household uses four 60-watt lamps and five 40-watt lamps for 5 hours a day. Find the cost of lighting in a month of 30 days if electric energy is charged 18 nP per B.O.T. unit. What is the maximum current drawn if the supply is at 220 volts ?

Solution : Total power $=4\times60+5\times40=440$ watts.
Energy consumed in a day $=440\times5$ watt-hours
,, ,, month $=440\times5\times30$,,
$=66000$ watt-hours $=66$ **kwh** or B.O.T. units.
Cost $=66\times18$ nP $=$ Rs 11.88.
Maximum current $=$ Watts/Volts $=440/220=2$ amp.

(2) A household uses two fans of wattage 110 each, one fan of wattage 60 and lamps of total wattage 380. The main fuse is one of 5 amp. The main voltage is 220. The tenant wants to use an electric stove of wattage 750. Will it be safe to use the stove when all the fans and lights are running ?

Solution : Total wattage of fans and lights
$=110+110+60+380=660$.
Current drawn when these run together $=660/220=3$ amp.
Current drawn by the stove $=750/220=3.4$ amp nearly.
Total current when all run together $=3+3.4=6.4$ amp.
Since the main fuse is one of 5 amp, it will be unwise to operate the stove along with the rest.

314. Thermocouple

When two dissimilar metals are connected together to form a closed circuit and one of the junctions maintained at a higher temperature than the other, a current flows round the circuit, Such a device is known as a *thermocouple*. If the metals are copper and iron, the current will flow from the copper to the iron across the hot junction (Fig. 352).

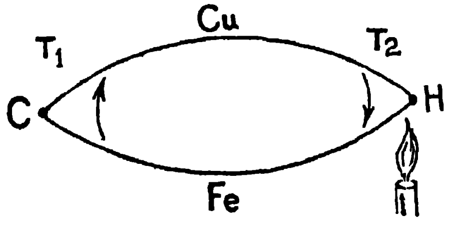


Fig. 352

The insertion of a third metal in the circuit will not alter the e.m.f. acting round the circuit if both junctions of the third metal are at the same temperature. This makes it possible to connect (Fig. 353) measuring devices, such as voltmeters, galvanometers etc. with the thermocouple without affecting its e.m.f.

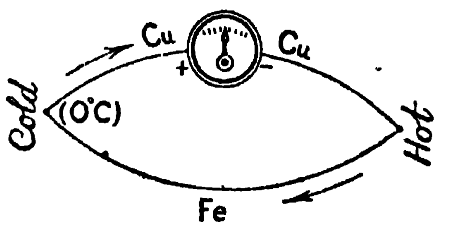


Fig. 353

The junctions of the metals are the seat of the e.m.f., though a very small part is contributed by the temperature gradient along the metals. The e.m.f. is always directed across a junction from one metal to the other. This e.m.f. depends on the temperature and generally increases with the latter.

When one junction of a thermocouple is kept at a constant temperature, say, that of melting ice, and the other gradually heated, the e.m.f. acting round the circuit increases. The relation between e.m.f. e and the difference t between the temperatures of the junctions is given by

$$e = at + bt^2$$

where a and b are constants for a given thermocouple. They can be found by measuring e for two known values of t.

Thermocouples are used extensively for the measurement of temperature. Different couples are useful over different ranges.

Thermocouple	*Range*
Copper—Constantan (60 Cu, 40 Ni)	$-100°C$ to $500°C$
Iron—Constantan	$-100°C$ to $800°C$
Chromel (10Cr, 90 Ni)—Alumel (98 Al, 2 Ni)	$0°C$ to $1100°C$
Platinum—Platinrhodium (90 Pt, 10 Rh)	$600°C$ to $1600°C$.

Direct reading thermoelectric thermometers (i.e. thermocouples) are used in industry for measuring temperatures of furnaces etc. Since a thermocouple can be made of very fine wires, the heat capacity of the junction will be very small. Such thermocouples can be used to measure temperatures of very small bodies, A thermocouple may be looked upon as a heat engine which converts heat energy into electrical energy.

Exercises

1. State Joule's law of heating and describe experiments to verify it.

2. How can the mechanical equivalent of heat be determined by electrical means ?

3. Define the following :

Joule, Watt, Watt-hour, Kilowatt-hour, B.O.T. unit.

A current of 10 amperes passes through a resistance of 20 ohms for 5 min. Find the power wasted as heat and also the total heat developed. (*Ans* : 2 kw; 1.43×10^5 cal.)

4. Mention some useful applications of the heating effect of currents for (*a*) domestic and (*b*) industrial purposes.

What is a fuse ? How does it act ?

5. Give a short description of the modern electric filament lamp. What do you understand by the following ?

(*a*) A 220-volt, 100-watt lamp.
(*b*) The efficiency of a lamp is two candles per watt.
(*c*) A coiled-coil lamp.

6. A current of 3·5 amp. flows for 10 min. through a coil of resistance 0.75 ohm. It is immersed in 200 gm. of water initially at 30°*C*. If all the heat is absorbed by the water, what will be the final temperature ? (*Ans* : 36·6°*C*.)

7. An 800-watt electric kettle has an efficiency of 80%. How long will it take to heat 1 kg of water from 30°*C* to 100°*C*? $J = 4{\cdot}2$ joules per calorie. (*Ans* : 7.7 min nearly.)

8. What is the current drawn by a lamp rated at 230-volts, 60-watts ? What is its resistance in use ? Find the cost of using this lamp for a month of 30 days at 4 hours a day when electricity costs 25 nP per unit. (*Ans*: ·26 amp ; 882 ohms nearly ; Rs. 1·80.)

9. Two lamps of resistances 500 and 200 ohms are connected across 100-volt mains. Which would glow more brightly ? Give reasons.

10. Two wires are of the same material and length ; but the diameter of one is half that of the other. The currents through them are so adjusted that the rate of production of heat is the same in both. Compare the currents.

(*Ans* : Current in the thin wire is half that in the thick one.)

CHAPTER 37

CHEMICAL EFFECTS OF CURRENTS

315. Some Terms and their Meanings

The passage of an electric current through a solution of an acid, salt or base in water is accompanied by chemical changes. It will be useful to know clearly the meanings of some terms in this connexion.

Ion. An ion is an atom, molecule or a radical (i.e., a group of atoms forming part of a molecule) which has an excess or a deficit of electrons over its normal quota. When there is a deficit of electrons, the ion has an excess of positive charge and is called a *positive ion.* A *negative ion* has an excess of electrons. (See § 312). for examples.)

Electrolyte. An electrolyte is a liquid in which electricity is conducted by means of both positive and negative ions. Solutions of acids, bases and salts in water are electrolytes. A solution of sugar is not. Oils and pure water are non-conductors of electricity.

Anode. The anode is the electrode or conductor by which positive electricity from an external source of e.m.f., such as a cell or a battery, is supposed to enter an electrolyte.

Cathode. It is the electrode by which positive electricity is supposed to leave the electrolyte.

Electrolysis is the chemical action that takes place at the interface between an electrode and an electrolyte when a current passes from one to the other.

A vessel containing an electrolyte with two electrodes of the same material introduced into it is called an **electrolytic cell.** In such a cell electrical energy supplied from an external source produces a chemical change. If, however, the electrodes are different in nature, the cell acts as a **voltaic cell** supplying electrical energy at the cost of chemical energy of the reactions that take place at the interfaces of the electrodes and the electrolyte.

316. Conduction of Electricity through Electrolytes

In electrolytes, the process of solution splits up or *dissociates* the molecules of the solute into two oppositely charged ions. The HCl molecule splits into a positive hydrogen ion, H^+, and a negative chlorine ion, $Cl,^-$ one electron having been transferred from the former to the later. The H_2SO_4 molecule splits into two H^+ ions and one SO_4^{--} ion, the double sign indicating that the SO_4 radical has an excess of two electrons. $CuSO_4$ splits into Cu^{++} and SO_4^{--}, $AgNO_3$ into Ag^+ and NO_3^-, NaOH into Na^+ and OH^- etc.

The chemical properties of an ion are very diffierent from those of the corresponding neutral atom or molecule. The ion is chemically inert. The Na^+ ion is stable in aqueous solution; but the neutral sodium atom, Na, will react vigorously with water.

When the electrodes of an electrolytic cell are connected to a battery, the negative ions are attracted towards the positive electrode, and the positive ions towards the negative electrode. Within the cell there are thus two streams of oppositely charged ions moving in opposite directions. These streams constitute the electric current in the electrolyte.

317. Some Electrolytic Reactions

Electrolysis of hydrochloric acid. Let a current be passed through a dilute solution of HCl in water, the electrodes being of carbon. Each H^+ ion, on reaching the cathode (i.e., the negative electrode), will combine with an electron and from a hydrogen atom. Such atoms will form molecules and hydrogen will be liberated as a gas from the cathode. At the anode, the chlorine ion, Cl^-, gives up its negative charge and changes to the chlorine atom. The liberated chlorine reacts with water and releases oxygen. Water, in the form of hydrogen and oxygen, is removed from the solution by electrolysis. not HCl.

Electrolysis of dilute sulphuric acid. If the electrodes are of such inert materials as carbon or platinum, the hydrogen ion, H^+, drawn to the cathode, behaves as in the previous case. Hydrogen gas is evolved at the cathode. The SO_4^{--} ion moves to the anode, gives up its charge, reacts with water and liberates oxygen. It is not H_2SO_4 which is removed from the solution by electrolysis, but water.

Electrolysis of water. An apparatus for the electrolysis of water is shown in fig. 354. If is filled with dilute sulphuric acid. The electrodes are of platinum. The liberated gases collect in the tubes below the stopcocks. The displaced liquid is driven up the central tube. The volumes of hydrogen to oxygen will be as 2 : 1.

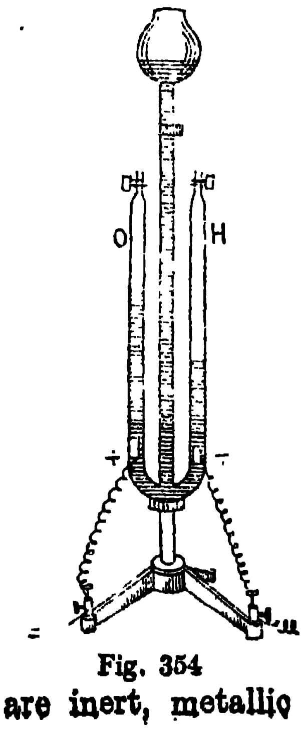


Fig. 354

Electrolysis of copper sulphate. The ions are Cu^{++} and SO_4^{--}. If the electrodes are inert, metallic copper will be deposited on the cathode. At the anode the sulphion, SO_4^{--}, will cause liberation of oxygen as in the electrolysis of water. Copper will be gradually removed from the solution.

If the electrodes are of copper, copper will be deposited on the cathode. But at the anode the sulphion will attack copper, which will dissolve and enter the solution as Cu^{++}. The strengths of the copper sulphate solution will remain unchanged. The net result of electrolysis will be a transfer of copper from anode to cathode. Electrolytic copper, which has a very high degree of purity, is industrially processed in this way out of impure copper ingots.

When silver nitrate solution is electrolysed with a silver anode, silver dissolves at the anode and forms a pure deposit on the cathode.

Fused lead chloride may be electrolysed into lead and chlorine by using carbon electrodes. Note that it is not a solution.

An interesting demonstration may be arranged by electrolysing lead acetate solution between lead electrodes. Lead crystals grow in the the form of a tree at the cathode, which should be at the centre (Fig. 355). At the anode the lead dissolves under the action of the acetate ion and keeps the strength of the solution constant.

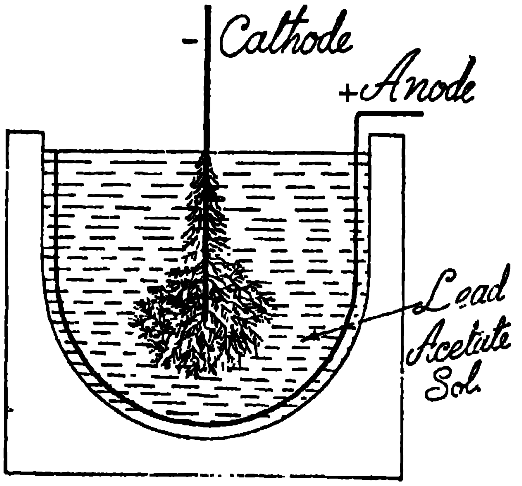


Fig. 355

Products of electrolytic decomposition. The above experiments show that

(i) the products electrolytic decomposition are not necessarily those the ions of which carry the current through the electrolyte. The products are the final results of the chemical actions taking place at the electrodes ;

(ii) metals and hydrogen are liberated at the cathode ; the acid radical or oxygen is liberated at the anode.

Determination of the polarity of a cell or the supply mains. Add a little starch paste to an aqueous solution of potassium iodide and moisten a piece of filter paper with it. If wires from the terminals of a cell are allowed to touch the paper a feeble current will pass and iodine will be liberated at the anode. Iodine reacts with starch to produce a dark blue coloration. Hence *the wire at which the dark blue colour is produced is the anode.* While testing the supply mains in this way, a suitable resistance, say, a lamp, must be included in the circuit to avoid risk of a short circuit.

When wires from supply mains are dipped into water to which a few drops of an acid has been added, gas evolves *more copiously at the cathode* than at the anode. This is so because hydrogen is liberated at the cathode, and the volume of hydrogen liberated in a given time is double that of oxygen. Here also a protective resistance should be included.

318. Faraday's Laws of Electrolysis

Faraday was the first to investigate systematically the conduction of electricity through electrolytes. By a series of brilliant experiments he was able in 1834 to formulate the following laws known as **Faraday's laws of electrotysis** :

I. The mass of a substance liberated by electrolysis is proportional to the quantity of electricity that passes through the electrolyte.

II. When the same quantity of electricity passes through different electrolytes, the masses of the various substances liberated are proportional to their chemical equivalents.

[The *chemical equivalent* of an element is its atomic weight divided by the valency. For a radical, it is the weight of the radical (on the chemical scale, i.e., taking oxygen as 16) divided by its valency.]

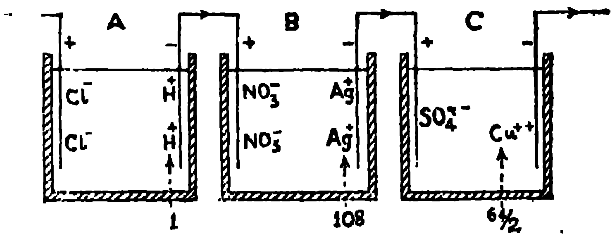


Fig. 356

Fig. 356 shows three electrolytic cells in series containing solutions of HCl, $AgNO_3$ and $CuSO_4$ respectively. When a current passes through the cells, it will have the same value in all of them as they are in series. H, Ag and Cu will be deposited at the respective cathodes. Their chemical equivalents are 1, 108 and 32 respectively. The masses deposited in a given time by a current will be in the ratio of these numbers.

If M is the mass of a substance liberated by a current I flowing for time t through an electrolyte, the charge that passes is $Q=It$. According to Faraday's first law of electrolysis M is proportional to Q, i.e., to It.

$$\therefore \quad M=ZIt$$

where Z is a constant known as the **electrochemical equivalent** (abbreviated E.C.E.). We may say that *the E.C.E. of a substance is the mass in grams of the substance liberated in electrolysis by the passage of one coulomb of electricity through an electrolyte containing the substance. It is expressed in grams per coulomb.*

Examples (i) How long will it take a current of 0·5 amp. to deposit 1 gm of copper. (E.C.E. of copper $=\cdot000329$ gm/coulomb.)

Solution : From relation $M=ZIt$, we have $t=M/ZI=1/\cdot000329 \times \cdot5=6078$ sec $=101\cdot3$ min.

(2) If 0·4752 gm. of nickel is deposited by a current of 0·6 amp. in 40 min, what is the E.C.E. of nickel ?

Solution : From relation $M=ZIt$, we have

$Z=M/It=0\cdot4752/\cdot6\times2400=3.3\times10^{-4}$ gm/coulomb.

Voltameters. By virtue of Faraday's first law of electrolysis it is possible to measure a charge by the mass of ion it deposits. Such devices are called *voltameters* (sometimes *coulombmeters* or *coulometers*). They are generally of simple construction. A copper voltameter has electrodes of copper placed in a strong solution of copper sulphate. Charge is measured by the increase in weight of the cathode. When the time of flow is known the value of the current is easily obtained.

319. Verification of Faraday's Laws of Electrolysis

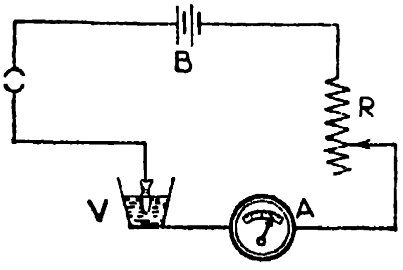


Fig. 357

The first law can be verified by using any voltameter. (Fig. 357 shows the circuit. The deposit for a known I and t is measured. I and t are altered in value in successive experiments and the corresponding deposit weighed. It wil be seen that the mass of the deposit is in every case proportional to $I\times t$, i.e., to the charge.

To verify the second law two voltameters, one having copper electrodes in $CuSO_4$ solution and the other, silver electrodes in $AgNO_3$ solution may be connected in series and a current passed through both. The masses of copper and silver deposited at the respective cathodes will be found to be proportional to their chemical equivalents.

320. Applications of Electrolysis

Electrolysis finds application in various fields. Some of the more important are (*i*) electroplating, (*ii*) electrotyping, (*iii*) extraction and refining of metals and (*ii*) manufacture of chemicals.

Electroplating. In electroplating, an article made of a metal is given a thin coating of a more attractive or durable metal by electrolysis. The articles to be electroplated are cleansed thoroughly and made the cathode of an electrolytic cell. The anode is made of the metal to be deposited. The electrolyte is a suitable solution which contains ions of this metal. To ensure an even deposit the cathode may be slowly rotated or be completely surrounded by the anode. Worn out machinery parts can be replated and brought to the right size by electrolysis.

In the table below, the first column gives the name of the metal used for coating articles made of metals named in the second column. The purpose of deposit is given in the third column.

Metal deposited	*Metal coated*	*Purpose of coating*
Cadmium	Iron	Rust-proofing.
Chromium	Brass	Untarnishable and unscratchable coating for car-parts, domestic fittings etc.
Gold	Brass	Imitation jewellery.
Copper	Iron	As foundation for nickel, silver or gold plating.
Nickel	Iron	Untarnishable coating for cycle parts.
Platinum	Brass	Hard, non-corrodible coating for scientific instruments.
Silver	Brass, Copper	Table utensils etc.
Zinc	Iron	Rust-proofing.

Electrotyping. Matter which has to be printed in large numbers is generally printed from *electrotype plates*. The matter is first set up in ordinary type and an impression is taken on wax. The wax mould is dusted with graphite powder and made the cathode in a bath of copper sulphate solution. The anode is of copper. When a current is passed, copper deposits on the mould which has been rendered conducting by the graphite. When the copper deposit acquires the thickness of a visiting card, it is peeled off the mould and backed up with type metal to give it necessary strength for printing.

Gramophone records are made by a similar process. The wax impression on which the vibrations of the recording needle are received is dusted with graphite and made the cathode in a plating bath for deposit of metal.

Extraction and refining of metals. Aluminium is extracted by electrolysing a fused mixture of alumina (aluminium oxide), cryolite and fluorspar. Only the alumina is decomposed. The electrodes are of carbon. The heat generated by the current keeps the electrolyte melted. As very large currents are necessary at a low cost, the industry is located near hydroelectric projects.

Copper, as it comes from the smelting works, is not very pure. To refine it the crude copper is made the anode and a sheet of pure copper the cathode of an electrolytic cell containing copper sulphate solution. When electrolysed, the deposited copper is is about 99·95% pure. It is known as electrolytic copper.

The metals sodium, potassium and calcium, the gas chlorine, and many other substances are now produced on a commercial scale by electrolysis. Chemicals like caustic soda, caustic potash etc. are produced electrolytically. Manufacture of chlorates and perchlorates in the explosives industry, of products of aniline by anodic oxidation, and of products of nitro-benzene by cathodic reduction illustrate a few of the many applications of electrolysis to industry.

Exercises

1. Explain the following terms :

Ion; electrolyte; anode ; cathode; electrolysis; electrolytic cell.

What is the difference between an electrolytic cell and a voltaic cell ?

2. Describe the actions that take place when

(*a*) Dilute HCl is electrolysed between carbon electrodes ;

(*b*) Copper sulphate is electrolysed between copper electrodes ;

(*c*) Copper sulphate is electrolysed between platinum electrodes.

3. State Faraday's laws of electrolysis. How would you verify them ? Define electrochemical equivalent and chemical equivalent.

4. Say how you would use a copper voltameter to measure a current. Draw the circuit and indicate clearly the positive and negative electrodes.

5. Mention some important applications of electrolysis.

6. How is a current conducted through an electrolyte ? Which of the following substances are electrolytes :

Paraffin oil, common salt solution, mercury, molten lead, sulphuric acid ?

7. 1·2 gm. of copper is deposited when a current of 1·5 amp. is passed through a copper sulphate solution for 40 min. Find the E.C.E. of copper. (*Ans* : ·00033 gm per coulomb.)

8. A Daniell cell gives a current of 0·1 amp. for 1 hour. Calculate the change in weight of both electrodes. (E.C.E. of copper =·00033 gm per coulomb. Chemical equivalent of copper= =31·8 ; of zinc=32·6).

(*Ans* : Zinc lost=·1218 gm ; copper deposited=·1188 gm.)

ELECTROMAGNETIC INDUCTION

321. Faraday's Discovery

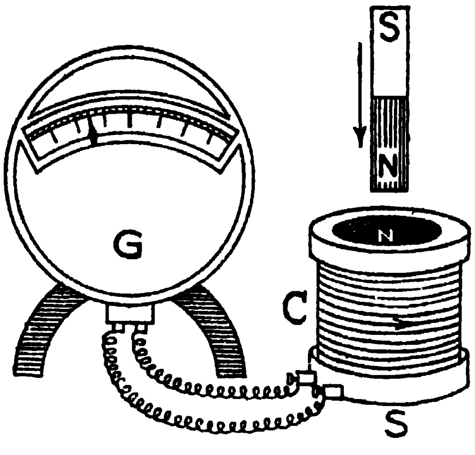


Fig. 358

Since Oersted's discovery in 1820 that a current always produces a magnetic field, physicists had been enquiring whether a magnetic field could produce a current. In 1831, Faraday made the discovery that a momentary current was produced in a coil of wire wound on an iron core when and *only when* another current was started or stopped in another coil on the same core. By a series of simple and well designed experiments he clarified the phenomenon which came to be known as 'electromagnetic induction'.

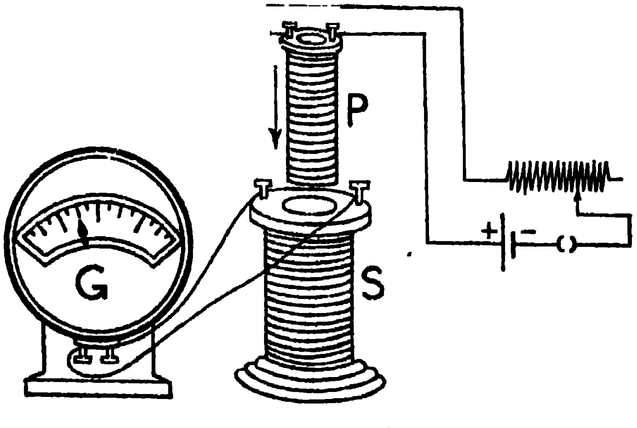


Fig. 359

The importance of this discovery in the development of modern industries and its effect on our social life cannot be overestimated. The electric generators which produce electrical energy and the transformers which efficiently transmit this energy across long distances, are direct applications of Faraday's discovery of electromagnetic induction.

322. Experiments on Induced Currents

Apparatus : To demonstrate the phenomenon of electromagnetic induction we may use the following apparatuses :

(*i*) A galvanometer capable of detecting small currents (*G* ; figures 358 and 359). It should be a centre-zero instrument.

(*ii*) A hollow coil of many turns of wire connected with the galvanometer (*S*; figures 358 and 359). One or more permanent magnets (*NS*; Fig. 358) or another narrow coil (*P*, Fig. 359) can be introduced into this coil. We should know in what direction a current must flow in the coil *S* so as to deflect the galvanometer needle to the right or the left.

(*iii*) One, preferably two, strong permanent magnets with their poles marked.

(*iv*) A narrow coil of many turns of wire connected to a battery (*P*; Fig. 359). A key and a rheostat are connected in series with the coil so that the current in it can be started, stopped, increased or decreased at will. This coil may be called the *primary coil* to distinguish it from the other coil which is called the *secondary coil.*

A. Currents induced by magnets. It is easy to show that whenever there is a relative motion between the magnet and the coil *S*, a current of short duration passes through the galvanometer. The current lasts so long as the relative motion lasts. When the magnet is moved slowly the galvanometer deflection is small. The deflection increases when the magnet is moved faster. It also increases when we use two magnets in place of one with their like poles together. These show that the current increases with (*i*) *the speed of the relative motion*, and (*ii*) *the strength of the magnet*. Suppose we know how the current must flow through the coil *s* so as to deflect the galvanometer to the right or the left. We can then find the direction of the current in the coil produced by a given motion of the magnet. The following observations can be made:

Relative motion between coil and magnet	*Direction of induced current in coil viewed from the side of the magnet*
N-pole approaching	Anticlockwise
,, receding	Clockwise
S-pole approaching	Clockwise
,, receding	Anticlockwise

Conclusions : The following conclusions can be drawn from the observations made above :

(*i*) The appearance of a transient current in the coil implies the appearance of a transient electromotive force acting round the coil.

(*ii*) This e.m.f. is present only when there is a relative motion between the magnet and the coil, i.e., when the magnetic field within the coil is changing with time.

(*iii*) The e.m.f. is stronger the quicker this change.

(*iv*) The e.m.f. increases when the area and number of turns of the coil is increased.

The production of an electromotive force round a circuit when a change occurs in the magnetic field embraced by the circuit, is called **electromagnetic induction.** The e.m.f. will appear even if the circuit is open ; there cannot be any current in the circuit in that case. The e.m.f. is called the **induced e.m.f.** and the current, the **induced current.**

B. Current induced by currents. Placing the primary coil *P* inside the secondary we can show that a transient current passes through the secondary when the current in the primary is *started, stopped, increased* or *reduced.* The direction of deflection of the galvanometer is the same when the current is started or increased. This direction is opposite to what we find when the current is stopped or reduced.

The primary coil behaves as a magnet so long as a current is flowing through it. Hence all the effects produced by permanent magnets can also be produced by the primary coil. The effects will be much magnified if a soft iron rod is placed within the primary.

323. Lenz's Law

When an induced current flows through a coil, the coil behaves as a magnet. From the direction of the induced current given in the table on the last page it will be seen that the polarity of the coil due to the induced current is such as to oppose the relative motion between the magnet and the coil. This observation may be expressed in the following form :

The direction of the induced e.m.f. is always such that, if it could produce a current, the magnetic action of the current would tend to resist the change by which the e.m.f. was produced.

This statement is known as **Lenz's law** of electromagnetic induction. It was discovered by Emil Lenz, a Russian scientist, in 1834.

It is easy to illustrate the application of Lenz's law in a simple case. Suppose a N-pole approaches a coil along its axis, as in the set up of fig. 358. By Lenz's law the magnetic effect of the induced current will tend to oppose this motion of approach. To do so the coil should present a north polarity to the approaching N-pole. Hence, looking from the side of the approaching N-pole, the current in the coil would appear to flow anticlockwise. Other cases of relative motion can be similarly treated.

Lenz's law and the principle of conservation of energy. We may want to know the source which supplies the energy of the induced current. When induction is due to relative motion, the agent which causes the motion does an extra amount of work against the repulsion or attraction between the induced current and the inducing magnet. This provides the energy of the induced current. When induction is due to a change of current in a stationary coil, extra work is done by the source of e.m.f. in the primary circuit.

Lenz's law can be derived from the principle of conservation of energy. If possible, let the direction of the induced current be opposite to what follows from Lenz's law. Then, when the north pole of a magnet is given a small displacement towards a closed coil, the induced current would attract the north pole instead of repelling it. This attraction would make the pole move towards the coil with increasing speed. Hence the induced current will grow stronger and stronger. The energy of motion or of the induced current is derived without an equal amount of work being done anywhere. This, according to the principle of conservation of energy, is impossible.

324. Flux Linkage and Electromagnetic Induction

All facts relating to electromagnetic induction can be conveniently described in terms of the concept of 'flux linkage'. 'Flux' in this expression stands for 'magnetic flux'. To understand what is meant by magnetic flux, let us think of the lines of force in a magnetic field. Imagine them to exist in such density that the number passing normally through unit area at a place is equal to the intensity of the field at that place. Consider now an area in a magnetic field. A number of lines of the above description will pass through the area. This number gives the magnetic flux through the area. When the area is normal to the lines, the flux through it is a maximum. If the plane of the area is parallel to the lines the flux is zero.

The lines passing through an area is the flux connected or linked with it. If the area is defined by the turns of a coil of wire, the flux linked with the coil is obtained by multiplying the flux through a single turn by the number of turns.

The facts of electromagnetic induction can be described in terms of magnetic flux in the following way:

(*i*) *Whenever there is a change in the magnetic flux linked with a circuit, an e.m.f. is induced in it.*

(*ii*) *The magnitude of the induced e.m.f. is proportional to the rate of change of magnetic flux linked with the circuit.*

(*iii*) *The direction of the induced e.m.f. is such that the current it drives tends, by its magnetic action, to keep the flux through the circuit unaltered.*

325. E.M.F. induced in a Conductor moving in a Magnetic Field

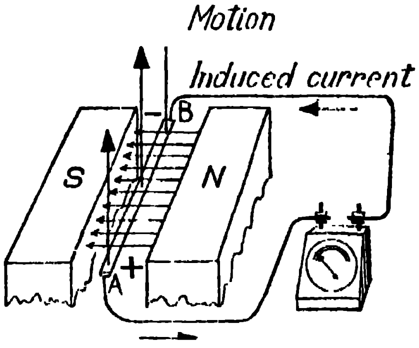


Fig. 360

When a conductor is moved in a magnetic field so as to cut across lines of force, an e.m.f. is induced between its terminals. To demonstrate it the ends of a metal rod *AB* (Fig. 360) may be connected by means of flexible wires to a sensitive galvanometer. If the rod is moved across the pole of an electromagnet, an induced current will be found to flow through the galvanometer. When moved parallel to the lines of force there is no such effect.

The direction of the induced e.m.f. is given by **Fleming's right-hand rule** (Fig. 361). It may be stated as follows:

If the thumb, fore-finger and middle finger of the right hand are stretched at right angles, the Fore-finger pointing to the direction of the magnetic Field and the thuMb to the Motion of the conductor, then the mIddle finger will give the direction of the Induced e.m.f.

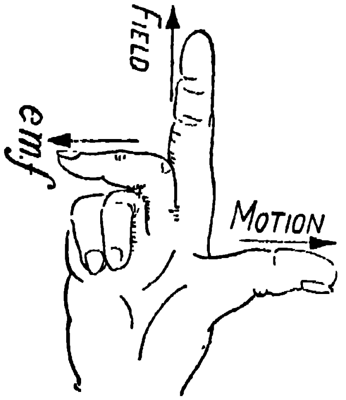


Fig. 361

We know that an e.m.f. is induced in a circuit or a coil when the magnetic field embraced by it (i.e. the magnetic flux linked with it) changes. An e.m.f. is also induced in a conductor moving in a magnetic field. These two are fundamentally the same phenomenon. *An e.m.f. is induced whenever there is relative motion between a magnetic field and a conductor.* It is immaterial whether the conductor moves or the field moves or changes. The total e.m.f. acting round a circuit is the sum of the e.m.f.'s acting along the conductors constituting the circuit.

It is easy to understand the action of a dynamo on the basis of the e.m.f. induced in a conductor as it moves in a magnetic field.

326. Introductory Remarks on A.C. and D.C. Generators and Motors

The electric **generator** is a machine which converts mechanical energy into electrical energy. It is usually driven by an engine or a turbine and feeds electrical power into any circuit connected to it. If the current set up by the generator in the external circuit is of an alternating type (i.e. if the current traverses the circuit first in one direction and then in the other, being reversed many times a second), the generator is called an **alternator,** *alternating current generator* or *a.c. generator.* If the current delivered to the outside circuit always flows in one direction, the generator is called a **direct-current generator**, *d.c. generator* or *continuous current* (*c.c.*) *generator.* There appears to be little uniformity in the use of the word *dynamo,* but more often it means a d.c. generator.

The electric **motor** (See § 304) is a machine which converts electrical energy into mechanical energy. It is a generator in reverse. When supplied with current from an electrical source it exerts a torque upon a shaft and turns wheels and machinery. The current supplied to the motor may be direct or alternating. In the former case, it is called a *d.c. motor*; in the latter, an *a.c. motor.* Motors can be constructed to run both on d.c. and a.c.; but they are generally of small power.

Principle of action of a generator. The action of a generator depends on the e.m.f. induced in a conductor when it cuts magnetic flux. The direction of this e.m.f. is given by Fleming's right-hand rule. The magnetic field is provided by magnets which, in commercial machines, are powerful electromagnets and are called **field magnets.** The conductors are placed in slots cut parallel to the axis on the surface of a laminated iron cylinder, called the **armature,** which rotates between the poles of the field magnets.

In large alternators the armature carrying the conductors is kept stationary and the field magnets made to rotate. To avoid confusion the rotating part of an alternator is called the *rotor* and the stationary part, the *stator.*

Principle of action of a motor. The action of a motor depends on the force exerted by a magnetic field on a conductor carrying a current. The magnetic field is provided by field magnets as in the generator. The conductors are also similarly arranged on an armature. Current from an external source is supplied to these conductors. The force on the conductors causes the armature to rotate. The relation between the directions of the current, field and motion is given by Fleming's left-hand rule. (See also § 304.)

327. The Simple Dynamo

A coil of wire so rotated in a magnetic field that the flux through it changes continuously constitutes a simple dynamo. Let us consider a rectangular coil placed in a uniform magnetic field (Fig. 362). Let it rotate about an axis perpendicular to the field but parallel to its sides *AB* and *CD*.

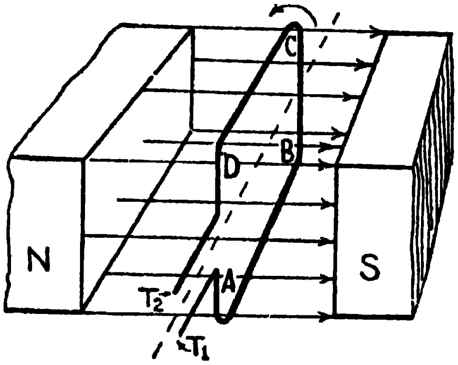


Fig. 362

As the coil rotates, the conductors *AB* and *CD* cut across lines of force. Application of Fleming's right-hand rule shows that

(*i*) The e.m.f.'s induced in the conductors act in the same direction round the coil.

(*ii*) The conductors *AD* and *BC* do not contribute to the e.m.f. as they do not cut across lines of force.

(*iii*) The terminals (T_1,T_2 ; fig. 362) of the coil become alternately positive and negative.

(*iv*) A terminal is positive during one half of a rotation, and negative in the next.

(*v*) The p.d. between the terminals is zero twice during each complete rotation, when the coil is perpendicular to the field. (The p.d. varies from zero to a maximum.)

To apply this e.m.f. to an external circuit the ends of the coil are connected to two slip-rings (R_1, R_2; fig. 363) mounted on the same shaft which carries the coil. The rings are insulated from each other. Two carbon brushes (B_1, B_2) press against the rings and are provided with a terminal each for connexion with an external circuit. As the coil rotates it drives an alternating current through the external circuit, and acts as an alternator.

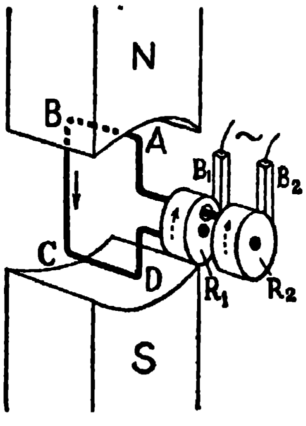


Fig. 363

328. The Simple Dynamo as a D.C. Generator

The simple dynamo described above can be converted into a simple d.c. generator by replacing the slip rings by a split metal tube or **commutator** as shown in fig. 364. The purpose of the commutator is to reverse the connexions of the coil with respect to the external circuit at the instants when the e.m.f. reverses in the coil. As the coil passes through its position perpendicular to the field, contact at each brush changes from one sector of the commutator to the other.

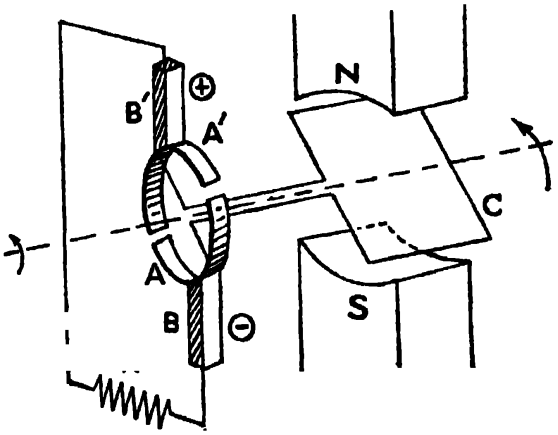


Fig. 364

The direction of the e.m.f. induced in a conductor moving across a given pole in a given direction is fixed. *The commutator ensures that such a conductor is always connected to the same brush.* This gives a brush the same polarity for all positions of the coil. Hence the current in the external circuit always flows in the same direction.

329. Mutual Induction and Self-induction

A. Mutual induction. A circuit carrying a current produces a magnetic field around it. If there is another circuit in the neighbourhood, a part of the magnetic field of the former will be embraced by the latter. In other words, some magnetic flux due to the former will be linked with the latter. If the current in the first circuit is altered in any way the flux through the second will change. Hence an e.m.f. will be induced in it according to the laws of electromagnetic induction. The phenomenon of the induction of an e.m.f. in a circuit (called the secondary) due to a change in the current through another circuit (called the primary) is known as *mutual induction*. The facts described in § 322 B relate to mutual induction.

The magnitude of the induced e.m.f. depends on (*i*) the rate of change of current in the primary, (*ii*) the number of turns of wire in each coil, and (*iii*) the extent of the magnetic field of the one embraced by the other. The induced e.m.f. increases as one or more of these quantities increase. A large enhancement of effect occurs if one or both coils have an iron core.

The transformer, which is one of the most important devices in applied electricity, depends for its action on mutual induction.

B. Self-induction. When a steady e.m.f. is applied to a circuit the current in it takes some time to grow to its full value. As soon as the current begins to flow, the magnetic field also begins to build up from zero value. The flux linked with the circuit therefore increases. It also increases when the current is increased. The flux diminishes when the current is switched off or reduced. In all such cases an e.m.f. is produced in accordance with the laws of electromagnetic induction. This phenomenon of the induction of an e.m.f. in a circuit due to a change in the value of the current in it is known as *self-induction.* The self-induced e.m.f. always opposes the growth and the decay of the current.

The magnitude of the self induced e.m.f. depends on and increases with (*i*) the number of turns in the coil, (*ii*) its diameter and (*iii*) the rate at which the current changes. Presence of iron in the coil greatly enhances the effect.

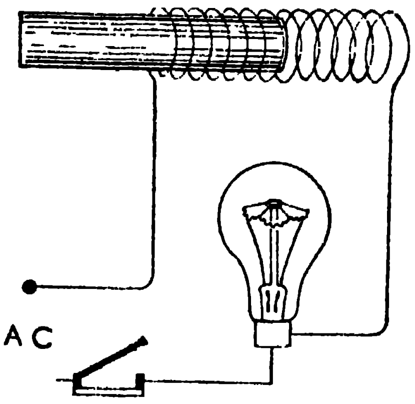


Fig. 365

The glow of an electric lamp fed by an alternating current can be reduced by introducing in its circuit a coil with an adjustable iron core (Fig. 365). As the core is introduced more and more into the coil the opposing e.m.f. due to self-induction increases. This reduces the glow of the lamp. The device is called a **variable choke.** It produces the same effect as a resistance, but wastes less energy. It can, however, be used only in alternating current circuits.

330. Transformers

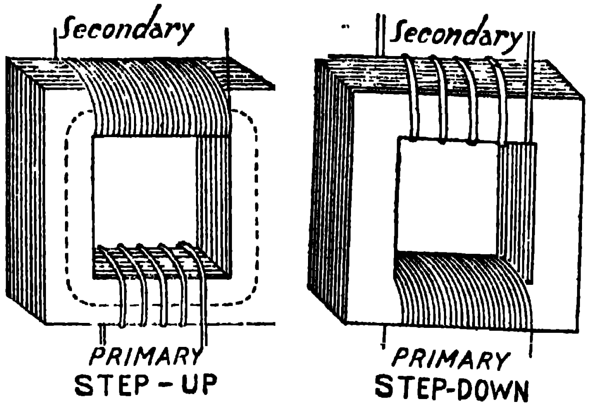


Fig. 366

A *transformer* is an electrical device by which electrical energy can be transferred from one circuit to another, generally with a change of voltage. It consists essentially of two coils, a primary and a secondary, wound upon a laminated and closed iron core (Fig. 366) and can be used only when the supply is alternating. The alternating potential difference is applied to the primary. The flux linked with a turn of the primary also passes through each turn of the secondary because the flux due to the primary is practically confined within the iron.

As the flux linked with the secondary changes due to the changing current in the primary, an e.m.f. is induced between the terminals of the secondary. In the ideal transformer (i.e., one in which no energy loss occurs and both coils embrace exactly the same flux), it can be shown that

$$\frac{\text{E.m.f. induced in secondary}}{\text{E.m.f. applied to primary}} = \frac{\text{Number of turns in secondary}}{\text{Number of turns in primary}}$$

The ratio of the secondary to the primary turns is called *transformation ratio.*

Neglecting losses we may write

Energy input in primary = Energy output in secondary

or power input = power output

or primary current × primary voltage

= secondary current × secondary voltage

$$\therefore \frac{\text{current in secondary}}{\text{current in primary}} = \frac{\text{number of turns in primary}}{\text{number of turns in secondary}}$$

Transformers may be *step-up* or *step-down* according as the voltage in the secondary is higher or lower than that in the primary. The current in the secondary of a step-down transformer is larger than that in the primary.

Economic transmission of electrical energy over long distances has been made possible by the transformer.

Whenever electric power has to be transmitted to a distance it is sent at as high a voltage as possible. This enables smaller currents, hence thinner wires, to be used, leading to a great economy in cable. Moreover, a smaller proportion of the energy transmitted is wasted by conversion into heat in the cable. Before supply to the consumer the voltage of the supply is reduced by step-down transformers. Smaller transformers are used in telephones, radio receivers and transmitters, television sets, in the production of X-rays etc.

331. The Induction Coil

The *induction coil,* also called the *Ruhmkorff coil,* is a device for obtaining a high voltage from a relatively low one, energy being supplied by a steady source of e.m.f. *In fact it is a transformer modified for use with direct current.*

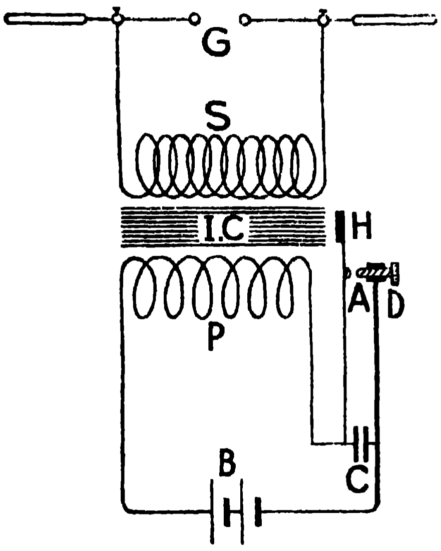


Fig. 367

Fig. 367 shows the essential parts of an induction coil. *P* is the primary coil ofwhich the windings consist of a relatively few turns of thick insulated wire wound round a core (*I.C.*) built up of soft iron wires. The secondary coil *S* consists of many turns of thin insulated wire wound over the primary. There are generally several layers, each well insulated from the other. The terminals of the secondary are connected to an adjustable spark gap *G*. For the sake of clarity *P*, *I.C*, and *S* have been shown separated in the diagram.

Since a steady current in the primary will not cause any induced current in the secondary, a make-and-break arrangement is provided in the primary circuit. It is generally a 'hammer break' working on the same principle as in the electric bell (See § 300). *H* is the soft iron armature which is responsible for the make and break at *A*. This occurs many times a second, and each make and break causes an induced current in the secondary. A condenser *C*, housed in the base of the apparatus, is joined in parallel with the make-and-break in the primary circuit. It enables a larger spark to be obtained across the spark gap at break.

Induction coils are used for running gas discharge tubes and X-ray tubes, and for setting up electric waves in radio telegraphy. The most important practical application of the induction coil is for ignition in gas engines. To make a spark jump across the terminals of the spark plug in such an engine several thousand volts are necessary. This is supplied by an induction coil.

332. Elementary Telephone

Electromagnetic induction made it possible to transmit the sound of human voice over long distances by electrical means. In the original telephone, designed by Alexander Graham Bell in 1876, transmitter and receiver were alike (Fig. 368). Two soft iron pole pieces (*P*) were

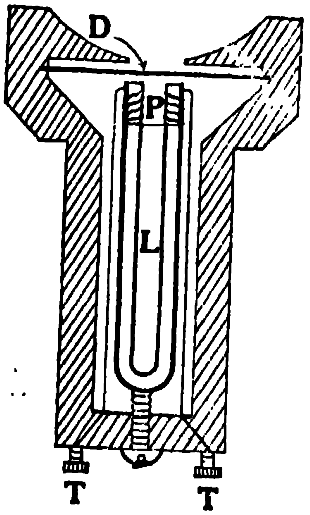


Fig. 368

attached to a permanent horse-shoe magnet (*L*). Round each pole piece a coil of very fine insulated wire was wound, the free ends of which were connected to terminals *T*. A disc *D* of thin sheet iron, secured at the edges, was placed very close to the pole pieces.

Sound waves falling on the disc caused it to vibrate in consonance with the waves. Its movements induced currents in the coils of the transmitter. These were carried along wires to the receiver which was exactly similar in construction. The induced currents, flowing through the coils of the receiver, altered the magnetization of the soft iron pieces. As a result the pull on the iron disc in the receiver varied, and it vibrated reproducing the movements of the disc of the transmitter and hence the sound which caused it.

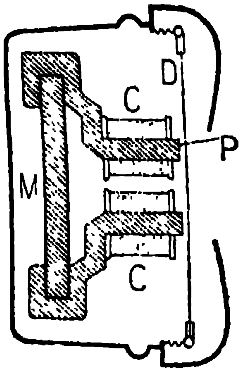


Fig. 369
P = Soft iron pole-pieces; *C* = Coils; *D* = Diaphragm of sheet iron; *M* = Permanent magnet.

The modern receiver is almost the same as that of Bell, but is more compact (Fig. 369). The modern transmitter is a carbon microphone (Fig. 370). It contains a little box filled with granules of carbon (*C*). The front (*B*) and back of the box are made of polished carbon plates. Current from a battery flows from one plate to the other through the granules. T_1, T_2 are terminals for connexion to the battery.

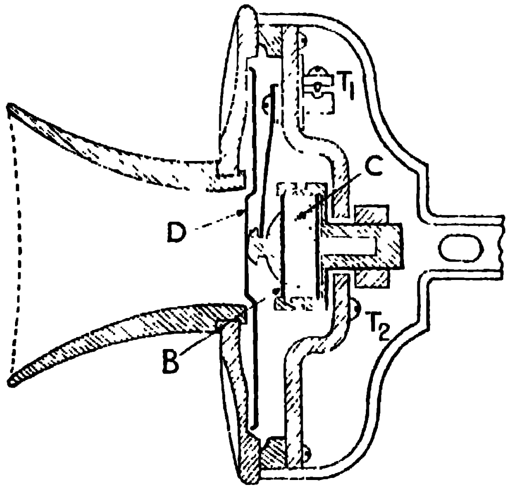


Fig. 370

B is attached to the centre of the diaphragm *D*, and moves in and out a little as *D* vibrates under the action sound of waves impinging on it. This alters the resistance in the circuit, since the resistance of the path through the granules depends on the tightness of packing. The current in the battery circuit, therefore, changes in tune with the sound vibrations impinging on *D*. This varying current actuates the coils in the receiver and causes its diaphragm to vibrate similarly.

A simple telephone circuit with a carbon microphone transmitter and a Bell receiver is shown in fig. 371. It can transmit a one-way message.

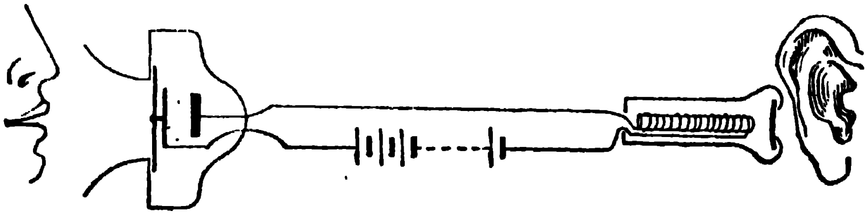


Fig. 371

333. Motor Generator and Rotary Converter

It is often required to use a potential difference of 110 or 220 volts when the supply in the mains is at a higher voltage. In such cases the transformation may be brought about by operating a motor from the mains, and making it drive a generator constructed to give the desired voltage. The motor and the generator are on the same shaft or on coupled shafts. A combination of a motor and a generator for the purpose of transforming any voltage, direct or alternating, to any other desired voltage within proper limits, is called a *motor generator*.

A *rotary converter* is a form of motor generator which consists of a combination of an alternating current motor and a direct current generator in which both motor and generator use the same armature windings. The alternating current is supplied through slip rings at one end of a rotating type armature. The direct current is carried by brushes from a commutator at the other end. These converters are used whenever a large amount of power, transmitted by alternating currents, is to be converted to direct current mains, as in the operation of electric trains or a d.c. town supply.

Exercises

1. What is electromagnetic induction ? Describe simple experiments to illustrate the phenomenon.

What is the most important law about the phenomenon ?

2. State Lenz's law of electromagnetic induction. Apply it to find the direction of the current induced in a coil when the south pole of a magnet is (*a*) brought near it, (*b*) taken away from it?

Whence does the induced current get its energy?

3. What do you understand by *magnetic flux* ? What is meant by *flux linked with a circuit* ?

State the principal facts of electromagnetic induction in terms of flux linkage.

4. How can you show that an e.m.f. is induced in a conductor when it moves in a magnetic field? State a rule connecting the direction of this e.m.f. with the directions of the magnetic field and the motion.

What will be the induced e.m.f. when a conductor moves parallel to a magnetic field?

5. Where do we apply Fleming's right-hand rule, and where the left-hand rule? How are they applied to the cases of the motor and the dynamo?

6. Describe with simple and neat sketches the action of a simple dynamo acting as an alternator.

How can it be converted into a d.c. generator? Explain the part the split-ring commutator plays in this regard.

7. What do you understand by mutual induction and self-induction? Describe simple experiments to illustrate them.

What applications do you know of these phenomena?

8. What is a transformer? How does it act ? Can you use it in a d.c. circuit? Why ?

9. What is an induction coil? How does it differ from a transformer ?

Draw a neat sketch of an induction coil and label its parts. Briefly describe its action.

10. Draw a sketch of a telephone receiver or transmitter and state how it acts. How should the two be connected for transmission of sound from one place to another?

CHAPTER 39

PRODUCTION, PROPAGATION AND DETECTION OF SOUND

334. Sound is due to Vibration of a Body

When a string in a musical instrument is plucked or bowed we hear a sound. That the sounding string is thrown into vibration is apparent to the naked eye. A metal pot when struck emits sound and its vibration can be felt by lightly touching it with the fingers. This vibratory movement is common to all objects emitting sound.

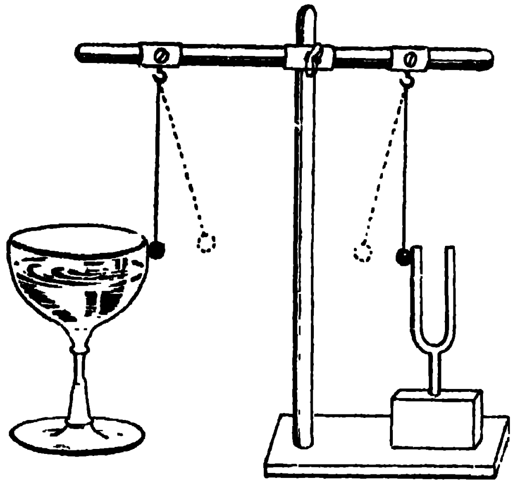


Fig. 372

That a sounding body is in vibration and has a certain amount of energy can be demonstrated by a simple experiment. Take a tuning fork (which is a U-shaped rod of steel provided with a handle) and strike it lightly. A sound will be emitted. Allow an arm of the tuning fork to touch a suspended pith ball. It will be thrown with a jerk and will move through a distance as a result of the impulse received from the fork (Fig. 372). The same phenomenon will occur if the pith ball is touched with any other sounding body, say a glass or a metal cup. When the vibration of any of the above bodies is stopped, the sound ceases. It is common experience that sound emitted by a vibratory body can be stopped by holding it firmly.

The vibrations produce waves in the surrounding air which, falling on the listening ears, produce the sensation of sound. The vibrating body is called the *source of sound* and the waves in the medium are called *sound waves.*

The periodic motion of the source of sound is the cause of sound. The frequency of the sound waves due to a source is the same as that of the source. If the frequency is lower than 30 or so or higher than 20,000 or so per sec, the waves cannot excite the human ear and no sound is heard. The waves generated by vibrations faster than the upper limit are called *supersonic waves.* Those slower than the lower limit are called *subsonic waves.*

Sound energy is no special form of energy. It is the mechanical energy of the vibrating source.

335. A Medium is necessary for the Propagation of Sound

Suspend an electric bell through a rubber stopper in a large bell jar (Fig. 373) placed on the receiver of an air pump. If an electric current is passed through the bell, its sound can easily be heard from outside. As air is gradually pumped out of the bell jar, the sound becomes fainter and fainter and then almost ceases. The hammer may be seen to strike the bell, but still no sound is audible. If air is now allowed to enter the jar gradually the sound becomes louder and louder and finally regains its initial strength. This proves that sound cannot be propagated in the absence of a material medium.

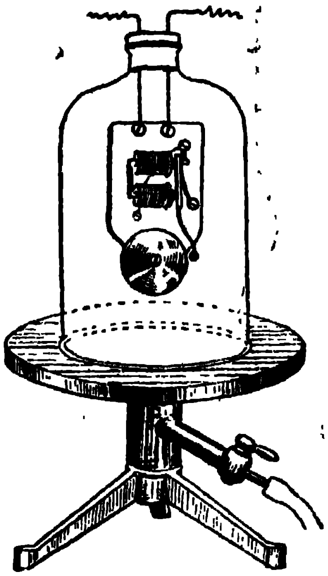


Fig. 373

Though the sound becomes very feeble it may not altogether vanish. A person standing close to the jar may yet hear a feeble sound. This sound is conducted along the suspending string and wires leading the current as well as through the residual air in the jar.

The following facts illustrate the propagation of sound through solids. If a watch is kept at one end of a table its sound may not be audible in air at the other end. But if the listener presses his ear to the other end, he may hear the sound propagated through the wood. The sound of an approaching train may be heard by pressing the ear on a rail before it is heard in air. American Indians are reported to have the ability to detect and even to locate the approach of enemy horsemen by pressing the ear to the ground.

A submarine's propellers produce sound waves in water which can be detected by a microphone placed in water.

336. Transverse and and Longitudinal Waves

Waves are ordinarily classified into two types, viz., *transverse* and *longitudinal*, according to the direction of vibration of a particle of the medium relative to the line of wave propagation.

A. Transverse Waves. When a stone is dropped into a pool of water a disturbance spreads out along the water in the form of circular crests and troughs of increasing radius. A piece of cork generated in one second. T is the time in which the wave advances by a distance equal to one wavelength.

Example. A tuning fork vibrating in air produces waves of wavelength $1\frac{1}{2}$ metre. Assuming the velocity of the waves so produced to be 330 metres/sec, calculate the frequency of the sound emitted and also its periodic time.

Solution. From $V = n\lambda$, $n = \frac{V}{\lambda} = \frac{330 \text{ metres/sec}}{1\frac{1}{2} \text{ metres}} = 220$ per sec

Therefore the frequency is 220 vibrations per sec.

The periodic time $= \frac{1}{220}$ sec $\left(\because T = \frac{1}{n}\right)$.

337. Mechanism of Sound Propagation

We shall now discuss how a vibrating source gives rise to longitudinal waves in air and how these waves are propagated.

Consider a vibrating tuning fork. As it vibrates, its arms move alternately inwards and outwards. When an arm moves from B to A (Fig.376) it compresses the layer of air in front of it. The particles of air in the compressed layer tend to relieve the strain of compression by moving to the right and compressing the adjoining layer. The particles of this layer in their turn compress the next layer and so on. In this way the compression advances to the right from layer to layer. By the time the arm reaches A this compression advances a certain distance depending on the time in which the arm moved from B to A. It also de-

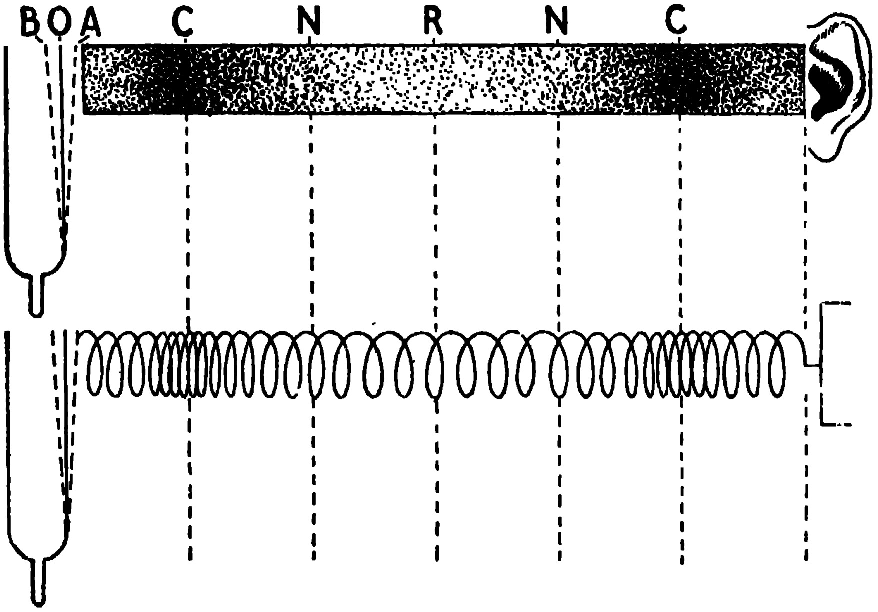


Fig. 376

pends upon the density and elasticity of the medium. As the arm swings back towards *B* it creates a partial vacuum behind it and the layer of air in contact with it expands. As the particles of this layer move to the left into this expanded region, they create a partial vacuum behind them, i.e. to the right. This is filled up by the particles from the next layer. In this way a rarefaction moves to the right and follows the compression moving in front of it. A compression and a rarefaction together form a complete longitudinal wave.

As long as the fork vibrates, alternate compressions and rarefactions take place in the surrounding medium. This alternately compressed and rarefied condition of the medium is transmitted through it and constitutes the sound waves. When they reach a listener's ear they produce alternately an increase and decrease of pressure on the ear drum. This causes nervous impulses to be carried to the brain, which interprets them as sound.

If a spring is attached to one arm of a fork as shown in fig. 376 alternate compressions and rarefactions may be seen along it when the fork vibrates. These can be distinguished if the movements are not very fast. A turn of the spring may be likened to a layer of air.

A vibrating body generates sound waves in the surrounding air as shown in fig. 377. A sensation of sound is produced if the frequency of the waves is within the appropriate range. Waves beyond this range do not produce the sensation of sound but they yet have the same physical properties. Hence they also should be called sound waves.

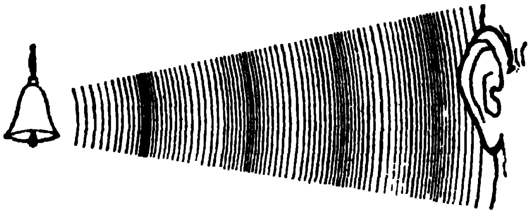


Fig. 377

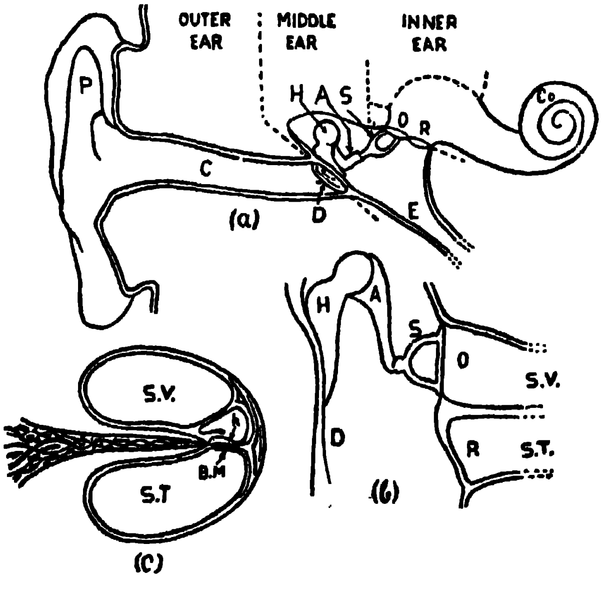


Fig. 378

338. Detector of Sound—the Human Ear

The **human ear** has three parts, viz., (*i*) the outer ear, (*ii*) the middle ear and (*iii*) the inner ear, as shown in fig. 378.

The outer ear. It consists of the visible part or the *pinna* (*P*), the *canal* (*C*) and the the *eardrum* (*D*). The pinna is a plate of elastic cartilage which catches the sound waves and conducts them through the canal to the ear drum or *tympanic membrane* which closes the inner end of the canal. The eardrum vibrates under the impact of the incident sound waves.

The middle ear. It is a cavity surrounded by bones. At one end of it lies the ear-drum and a membrane (*O*) at the other. This membrane is called the *Oval Window* or *Fenestra Ovalis* and forms part of the wall of the inner ear. *D* and *O* are connected by a chain of three small bones forming a bridge. They are respectively known as the *hammer* (*H*), *anvil* (*A*) and *stirrup* (*S*) because of structural similarity. These bones act as levers and transmit the vibration of the ear-drum to the liquid in the inner ear. They reduce the movement by about 30 to 1 and step up the pressure. The stirrup closes the aperture of the inner ear. To maintain equal air pressure on both sides of the ear-drum, the middle ear is connected to the throat by the *Eustachian tube* (*E*). Fig. 378 (*b*) shows the middle ear and a part of the inner one.

The Inner ear consists of a bony tube about an inch and a half long coiled into a spiral and called the *cochlea* (*Co*). It forms the actual hearing chamber and is divided throughout into two parts by a membrane stretched along the middle. On this membrane there are several thousand tiny fibres of varying lengths and tensions somewhat like the strings of a harp. They are connected by nerves with the brain. The cochlea is filled with a liquid. A transverse section of the inner ear is shown at (*c*) in fig. 378. *S. V* and *S. T* are the two chambers and *B. M* the membrane separating them. *h* represents a fibre.

Action. Sound waves entering the canal of the ear give rise to vibrations of the ear-drum. These vibrations are transmitted through the hammer, anvil and stirrup to the liquid in the cochlea and excite responses in some of the fibres. Nervous impulses are thereby set up and carried along the auditory nerve to the brain where they are interpreted as sound. These impulses are electrical in nature ; how the brain interprets them is not known.

The ear is perhaps the most sensitive detector of sound. "It can interpret as sound a vibration of amplitude 13×10^{-8} cm and register the reception of energy at the rate of 10^{-12} watt. If energy was passed into one cubic centimetre of water at this rate and converted into heat without loss, it would take 130,000 years to raise the temperature by 1°C."

339. Limits of Hearing

The slowest vibrations which the human ear can recognize as sound, have a frequency of about 30 per sec. If the frequency is above a certain limit we cannot hear it at all. This upper limit varies from man to man, the normal being about 20,000 vibrations per sec. With advancing age this limit is lowered.

Vibrations above the range of hearing are called **supersonic vibrations.** Many animals, including dogs, cats, bats and birds, can hear sounds above the upper limit of human hearing. Bats can fly in the dark avoiding all obstacles. It is believed that they produce supersonic sound as they fly. From sound reflected by the obstacles, bats detect their presence and position. Depths in sea can be determined by the reflection of supersonic vibrations from the sea-bed (See § 343).

340. Sound travels with a Finite Speed

It takes a finite time for sound to travel from the source to the listener. A flash of lightning and the accompanying peal of thunder are practically simultaneous events. Since light travels with the enormous speed of 186,000 miles per sec., the time it takes to reach a spectator at a distance of a few miles is negligible. But sound is not so fast. Hence we always hear the peal of thunder some time after we have seen the lightning flash. Similarly, when a gun is fired within visible distance, the report is heard a little after the smoke is seen. The interval between seeing and hearing in both the cases depends upon the distance of the cloud or the gun. Knowing the velocity of sound this distance can be calculated.

The speed of sound in dry air at $0°C$ is 1089 ft/sec or 331·4 metres/sec. It varies from medium to medium. In water the speed is about 1457 metres/sec., and in iron, about 5000 metres/sec.

341. Open Air Determination of Velocity of Sound

In 1864 Regnault determined the velocity of sound in air. A gun was fired from the top of a mountain and the corresponding flash of light was observed from another mountain some miles away. The observer measured the time interval between seeing the flash of light and hearing the report. If the distance between the stations was S and the time interval t, the velocity of sound would be $V=s/t$.

Sources of error. Though the arrangement gives the velocity of sound very easily, the method cannot yield accurate results because of the following reasons:

(*a*) It is impossible for anyone to start his watch at the instant that he sees the smoke. There is always an interval of time, called the *reaction time*, between the two actions. The error due to it diminishes as the distance separating the stations increases.

(*b*) The velocity of wind affects the velocity of sound. When both sound and wind move in the same direction the apparent velocity of sound increases. If they are in opposite directions, the apparent velocity diminishes.

To minimize the effect due to reaction time sensitive instruments are used for recording the moment the gun is fired and the moment the sound arrives at the distant station. To eliminate the effect of wind reciprocal observations are made, i.e., the sound is made at each station and its time of arrival at the other noted. The mean value of the two intervals is the time sound will take to travel the distance between the two stations in the absence of wind.

Factors affecting the velocity of sound in air:—

(*i*) The velocity of sound in air increases by about 61 cm/sec for every 1°*C* rise of temperature.

(*ii*) Sound travels faster in humid air than in dry air.

(*iii*) The apparent velocity of sound increases if both sound and wind move in the same direction.

(*iv*) The velocity of sound does not change with pressure only.

Table. *Velocity of sound in different media*

Gases at N.T.P.		*Solids and water*	
	metres/sec		metres/sec
Air	331·9	Copper	3600
Hydrogen	1286	Steel	5000
Oxygen	317·2	Glass	5100
Water-vapour	400	Marble	3800
Sulphur dioxide	209	Water (at 13°C)	1437

342. Reflection of Sound

Sound waves follow the same geometrical laws of reflection as light waves. Sound is reflected from walls, hills, rows of trees etc. One of the most important examples of reflection of sound is the echo.

Practical application of reflection. To transmit sound over not very long distances without loss of intensity a *speaking tube* (Fig. 379) is often used. If a sound is produced at the mouth of a tube, the sound waves that enter the tube

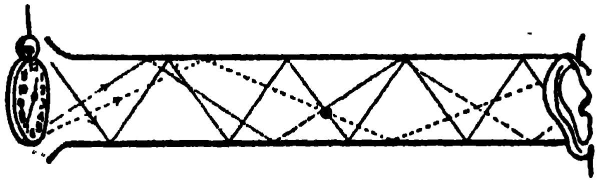


Fig. 379

cannot spread out in a rapidly increasing volume of air. They suffer repeated reflections at the inner walls and are confined to the air in the tube. For this reason, the intensity does not decrease according to the law of inverse square as it would do if the waves could spread out. The doctor's stethoscope works on this principle. The horn of a gramophone or a megaphone similarly prevents the spreading of sound and keeps the intensity high

The "whispering gallery" of St. Paul's Cathedral is circular in shape and a person talking quietly on one side can be heard on the other by multiple reflection of the sound waves round the wall.

343. Echoes

A sound heard by reflection and clearly distinguished from the original is called an **echo.** For its formation, an echo requires a suitable reflector such as an extended wall, a cliff etc., whose irregularities are small compared with the wave-length of the sound. To hear the sound and its echo *separately* the reflector must be at a sufficient distance, which may be found as follows :

A sensation of sound persists in the brain for almost 1/10 of a second after it is heard. Hence to recognise a sound and its echo as separate, at least this interval must elapse between them. In our country, the velocity of sound under ordinary conditions, (a mean temperature of 25°*C* and a fair amount of moisture in the air) is about 1150 ft/sec. In 1/10 of a second sound travels 115 ft in air. Thus, to hear distinctly the echo of a sound of *short duration*, the distance between the source and the reflector must be at least 57·5 ft, it being presumed that the source and the listener are close together.

The **rumbling and rolling of thunder** is due to echoing of a peal of thunder from a number of reflecting surfaces such as clouds, rocks, mountain sides, forests, surface of separation between air currents etc. The continuous rumbling is due to the fact that echoes reflected from different sources enter the ear at intervals less than 1/10 of a second.

In big halls it is often noticed that when a loud sound has ended a continuous *rolling* of sound persists for some time. This is known as **reverberation** and is due to multiple reflections from the walls. To prevent reverberations and enhance the absorption of extraneous sound, large halls and cinema houses are carpeted and part of the walls is thickly cushioned with soft sound absorbing materials.

Echoes are exploited by whalers in Arctic waters to locate an iceberg in a dense fog or at night. For this purpose a whistle is sounded or a short sharp sound made. By timing the interval between the signal and the arrival of the echo an iceberg may be roughly located and the boat steered clear of it.

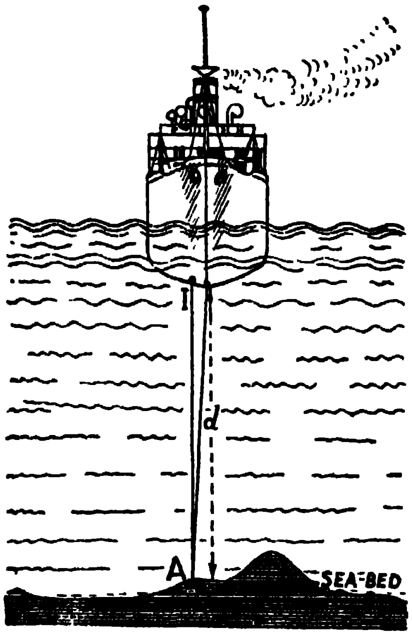


Fig. 380

Echo depth-sounding. Reflection of sound is employed as a means for finding depths at sea (Fig. 380). There are many devices for this purpose, of which the details differ. But the principle is as follows: A high frequency, generally supersonic, sound of short duration is produced near a hydrophone (a device similar to a microphone) submerged in water. The hydrophone responds twice, once to the original sound and again to the waves reflected from the sea-bed. The time interval between the two events is automatically recorded on a device which also measures short time intervals accurately. If the time between the two responses at the hydrophone is t and the velocity of sound in sea-water is V, then the depth at the place is $d=\frac{1}{2}Vt$.

Submarines can be detected when submerged in water by an apparatus called *Asdic* which works on the same principle as described above. It sends out pulses of very high frequency supersonic waves, which are reflected back to the apparatus by the submarine. Shoals of fish may be detected in the same way.

Examples. (1) A sound of very short duration is echoed back from a distant wall after 1/5 sec. If the velocity of sound is 330 metres/sec what is the distance of the wall from the source ?

Solution. It is clear that sound has travelled from the source to the wall and back in $\frac{1}{5}$ sec. It has therefore traversed a distance of $\frac{1}{5}\times 330=66$ metres. Hence the distance of the wall from the source$=66/2=33$ metres.

(2) A person standing between two parallel cliffs fires a gun. He hears the first echo after 3 sec and the second after 4 sec. Calculate the distance between the cliffs assuming the velocity of sound to be 1110 ft/sec at the place.

Solution. The first echo is received after 3 sec. Hence the sound has been reflected $1\frac{1}{2}$ sec after the firing. In that time sound travels $1110 \times 1\frac{1}{2} = 1665$ ft. Hence the distance of the nearer cliff is 1665 ft. Similarly, the distance of the other cliff from the observer is $1110 \times 2 = 2220$ ft. So the distance between the two cliffs is $1665 + 2220 = 3885$ ft.

344. Free Vibration

When a pendulum oscillates, it does so with a characteristic frequency determined by its length and the value of acceleration due to gravity at the place. A steel strip or a spring, suitably clamped, also vibrates with a characteristic frequency determined by its dimensions and elastic properties.

Vibrations in bodies may be excited by plucking, striking etc. After the exciting force has been removed the body continues to vibrate with a frequency characteristic of its own. Such a vibration is called a **free vibration.** The frequency with which it is executed is called the **natural frequency** of the body, and the corresponding periodic time is called the **natural period.** Free vibrations may be transverse or longitudinal.

345. Forced Vibration and Resonance

Let a periodic force (i.e., a force which attains the same value at equal intervals of time) be applied to a body capable of vibration. The body will at first tend to vibrate with its natural period. But after some time it will be found to vibrate with the period of the applied force. It will continue to do so as long as the applied force acts. Such a vibration of the body, i.e., one in which the periodicity is that of the applied force, is called **forced vibration.**

If the periodic time of the applied force is identical with the natural period of the body, a large vibration may be built up by the application of even a small force. This kind of forced vibration in which the period of the applied force and the natural period of the vibrating body are identical is known as **resonance** or *sympathetic vibration.* The amplitude is small in forced vibration, but large in resonance.

Demonstration of forced vibration and resonance. Fix a wooden stick horizontally and suspend from it a number of pendulums (Fig. 381) of which *A* should be heavier than the rest. The lengths of the pendulums *B*, *C*, *D* are different from that of *A*; but the lengths of *A* and *A′* are equal.

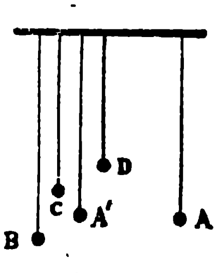


Fig. 381

If *A* is set oscillating at right angles to the stick, *A′* will soon be found to vibrate with a large amplitude. But the amplitudes of other pendulums will be small. The vibration of A exerts a periodic force on the stick. This is transmitted to the other pendulums and throws them into periodic vibration. The periodic time of the applied force agrees only with the natural period of A'. So A' is thrown into resonance or sympathetic vibration, while other pendulums undergo forced vibration of small amplitude.

Resonance and forced vibration in musical instruments. In stringed instruments like sitar, esraj etc., strings are stretched on a thin wooden board. The vibration of a wire produces forced vibration of the board and thence of air. This intensifies the emitted sound.

Each instrument has also several strings on it tuned to different notes. When a note is sounded on the principal wire, sympathetic vibrations are excited in the strings tuned to the same note. This increases both the intensity and the pleasantness of the tune.

346. Measurement of Wave-length by Resonance Tube

A wide glass tube, about two inches in diameter and 50 to 100 cm in length, may serve as a simple resonance tube. The other accessories are a tall glass jar filled with water, a stand to hold the tube and a tuning fork, as shown in fig. 382.

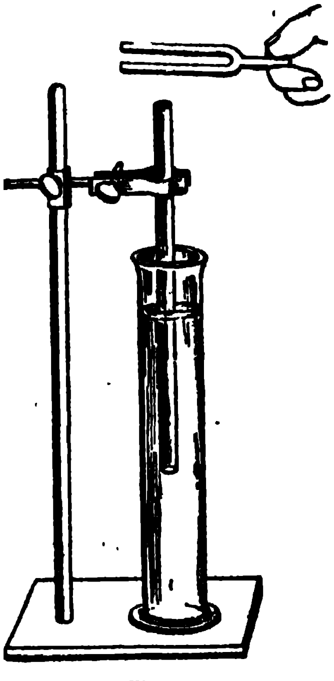


Fig. 382

This simple arrangement will enable us to find the wave-length of the sound emitted by the fork. For this purpose, sink the resonance tube to the maximum depth. Then slowly raise it and, as you do so, move the vibrating fork, also slowly, across the mouth of the tube. At some particular and well-defined position of the tube, the air within it will respond to the vibrations of the fork. A fairly loud sound will then be emitted by the tube. Adjust the height of the tube so that this sound is a maximum.

Measure the length of the tube above the water level; let it be l. Then $4l$ is approximately the wave-length (λ) of the sound emitted by the fork.

To get a more accurate result a correction, known as *end correction*, has to be applied (See § 356). It states that l should be increased by $0{\cdot}3d$, where d is the inner diameter of the resonance tube. Hence the wave-length is given by

$$\lambda = 4(l + 0{\cdot}3d)$$

Velocity of sound. If the frequency (n) of the fork is known, the velocity (V) of sound in the air of the resonance tube is given by

$$V=n\lambda=4n\ (l+0{\cdot}3d)$$

[The resonating length of a fork of frequency 128 will be about 62—63 cm at the average room temperture.]

347. Beats

When two sounds of nearly equal frequencies and loudness are sounded together, the loudness of the resulting sound rises and falls at regular intervals. The alternate rise and fall of sound, so brought about, is known as **beats.** A rise and a fall together constitute one beat. **The number of beats per second is equal to the difference in the frequencies of the two sounds.**

To demonstrate the phenomenon of beats take two large tuning forks of the same frequency. Load an arm of one of the forks with a small quantity of wax, or wind one or more turns of fine wire around it, near its free end. Loading an arm lowers the frequency of a fork. The two forks now have slightly different frequencies. If they are sounded together beats will be heard. If the load on the arm is increased, beats will be faster. The number of beats per second is equal to the difference of the frequencies of the two forks. It is difficult for an untrained ear to detect beats if there are more than five or so of them per second.

A load placed near the free end of an arm lowers the frequency of the fork more than an equal load placed elsewhere on it. This can be tested by shifting the load (of wax or turns of wire) along the arm and counting the number of beats it produces in a given time when sounded with the other fork.

Example. P is a fork of frequency 256 per sec. A fork Q produces 4 beats per sec when sounded with P. On loading Q slightly the number of beats changes to 2 per sec. What is the frequency of Q?

Solution : The frequency of Q differs from that of P by 4. We have to decide whether it is higher or lower than that of P. Loading Q lowers its frequency. The number of beats falls from 4 to 2 in this case. Clearly, the frequency of Q is higher than that of P. For if it were lower, further lowering of its frequency would have increased the number of beats.

Hence the required frequency$=256+4=260$ per sec.

348. Stationary Waves

Melde's Experiment. Take a large tuning fork of relatively low frequency, say, 64 vibrations per second. Fix it to a stand as shown in fig. 383. Attach one end of a string to one arm of the fork. Pass the string over a pulley and join its other end to

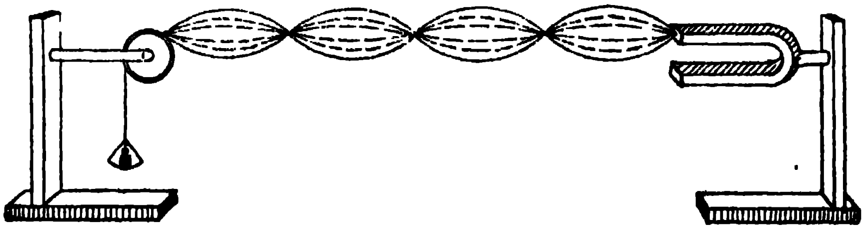


Fig. 383

a pan. Make sure that the fork vibrates perpendicular to the string.

Put some load on the pan and excite the fork. By adjusting the value of the load on the pan or the length of the string, you can make the whole string vibrate in one loop. If the load is reduced to one-fourth of its value, the string will vibrate in two loops. When it is reduced to one-ninth, three loops will be produced.

Stationary waves. The vibrations of the fork produce transverse vibrations at one end of the string. They move forward as waves along the string, and are reflected from the other end. The incident and reflected waves are superposed on each other producing the patterns of vibration that we see. These patterns are known as stationary waves. We may, therefore, say :

When two identical systems of waves, moving in opposite directions, are superposed, they give rise to patterns of vibration which are called **stationary waves.**

Stationary waves, also called *standing waves*, occur frequently in nature. Whenever waves are formed in a *limited medium*, reflection occurs at the boundary. These reflected waves are superposed on the original waves, and produce stationary waves.

Give a rap to a circular bowl containing water. You will find a stationary wave pattern on the surface of water.

Characteristics of stationary waves. Stationary waves in a medium (whether it is a solid, a liquid or a gas) have the following characteristics. It would be instructive to look for them in the vibrations of the string just described.

(*i*) The amplitude of vibration varies from place to place. Points of maximum amplitude are called **antinodes.**

(*ii*) At certain points the medium is at rest. Such points of no displacement are called **nodes.**

(*iii*) Along the common line of propagation of the waves, nodes and antinodes alternate with each other.

(*iv*) All particles between consecutive nodes move in the same direction at the same time. (The portion of the medium between consecutive nodes is called a *loop*.)

(*v*) Particles in neighbouring loops move in opposite directions.

Wave-length relations. The **wave-length** (λ) of a stationary wave is the same as the wave-length of the waves which, by their superposition, form the stationary wave. The **frequency** of a stationary wave is the number of times a loop of the wave vibrates in a second. It is also the frequency of the generator waves.

Distance between two consecutive nodes
=distance between two consecutive antinodes
=half the wave-length (λ/2).

Distance between a consecutive node and antinode
=a quarter wave-length (λ/4).

349. The Tuning Fork

The tuning fork is a simple piece of apparatus indispensable in the study of sound. A fork, properly designed and constructed, emits a pure tone when mildly struck. It is a bar or rod of steel bent into the form of a long U, and has a stem (or handle) at the bend. The arms of the U are generally called *prongs*. While vibrating, the prongs of the fork move towards and away from each other, as shown in fig. 384. There are two places near the bend (*N* and *N'*) which remain stationary. They are called *nodes*. The bend along with the stem is alternately depressed and elevated as the prongs vibrate inwards and outwards. The stem thus has an up and down motion. Forks of high frequency are short and thick. Loading a fork diminishes its frequency. Tuning forks are sometimes mounted on hollow wooden boxes with one side open. The dimensions of a box are so chosen that the air inside it has the same natural frequency of vibration as the fork. When the fork vibrates, the up and down motion of its stem throws the board on which it is mounted into forced vibration. This in turn excites resonance in the air inside. The sound emitted is louder than that due to the fork alone. The vibrations, however, die out more quickly.

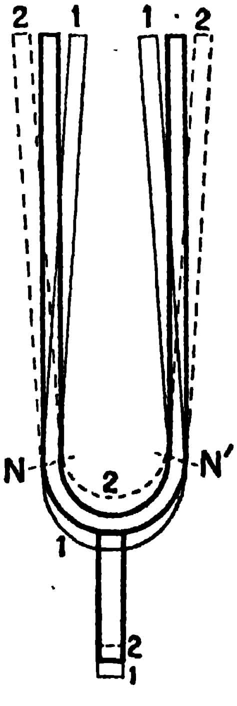


Fig. 384

A fork can be maintained in vibration by electrical means.

This requires an electromagnet, and a make-and-break device for the current, the device being attached to a prong of the fork. The arrangement has much in common with an electric bell.

350. Noise and Musical Sound

The sounds we hear may be divided into two classes, viz., (*i*) noise and (*ii*) musical sound. A *noise* is a sound of short duration or one which changes its character continually if it persists. The slamming of a door, the rumbling of a truck or a peal of thunder is an example of a noise. It is due to vibrations which have no regular periodic nature.

There is much noise in offices and factories. Noises reduce efficiency of workers by the disturbing effect they produce. The level of noise at different places of work has been measured and its effect on efficiency determined. Steps have also been devised to reduce noise in factories.

A *musical sound,* also called a *note*, is one which is due to vibrations of a regular periodic nature.

351. Characteristics of Musical Sound

Musical sounds have three important characteristics, viz., (*i*) loudness, (*ii*) pitch and (*iii*) quality.

(i) **Loudness.** Of two sounds similar in other respects, the one which can be heard from a longer distance is said to be the louder of the two. 'Loudness' in sound corresponds to 'brightness' in light.

Loudness and intensity. The amount of energy sound waves carry in one second across unit area is called the intensity of sound. Loudness, which is a physical sensation produced in the ear, seems to vary as the logarithm of the intensity. Loudness increases with (*a*) the size of the source, (*b*) the amplitude of vibration of the source and (*c*) the density of the medium carrying the sound. It diminishes with increasing distance.

(ii) **Pitch.** 'Pitch' corresponds to 'colour' in optics and refers to the sensation of sharpness or flatness which a sound produces. A noise has no definite pitch.

The pitch of a note is determined by the frequency of vibration of the source. As the frequency increases the note becomes shriller and is said to have a higher pitch. Though the word 'pitch' strictly refers to a sensation, it is often used in the same sense as 'frequency'.

(iii) **Quality.** The characteristic which differentiates between two notes of the same loudness and pitch, but produced by sources of different kinds, is called quality. There is no difficulty in distinguishing between a note sounded on a flute and the same note sounded on a sitar. The character by which we differentiate them is quality. It is by quality that we recognize the voice of a friend.

The physical character of the sound vibrations which determines the quality of a note may be understood as follows. A musical sound is seldom pure, i.e., it seldom consists of vibrations of only one frequency. The vibrations of the lowest frequency in a note is called the *fundamental*. Along with this there are notes of higher frequencies, which are called *overtones*. The number of overtones present in a note, their frequencies, and intensities relative to the fundamental, determine the quality of a note. Quality is also called *timbre* or *tone structure*.

352. Vibrations of Strings

When a stretched string, fixed at its ends, is plucked or struck, transverse waves are produced along it. These waves move towards the fixed ends and are reflected there. The outgoing and reflected waves produce stationary vibrations (§ 348) in the string. A string can vibrate in more than one way.

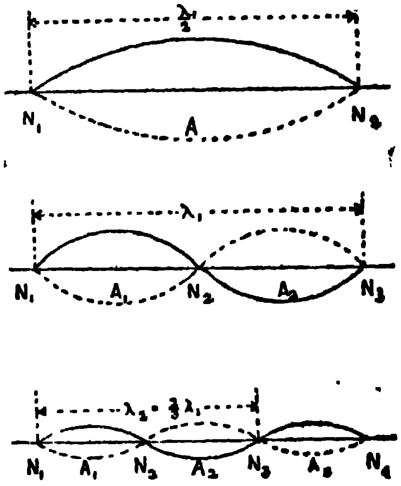


Fig. 385

Fundamental frequency. As the two ends of the string are fixed, they cannot vibrate. Hence two nodes (§ 348) of the stationary waves are always formed at these points. The simplest mode of vibration of the string is one in which these are the only nodes in the string. The antinode is then at the middle of the string. In this mode of vibration the entire string moves to and fro in a single loop (Fig. 385; top). The frequency of this vibration is called the *fundamental frequency*. Since the distance between two consecutive nodes is half the wave-length, the length of the string (L) will be related to the wave-length (λ_1) by the relation.

$$L = \lambda_1/2.$$

Harmonics A stretched string can also be made to vibrate with one or more additional nodes between those at the two ends. If there is only one extra node, it will be at the middle of the string

(Fig. 385; middle). The string will vibrate in two loops. The wave-length of the emitted note will be half that of the fundamental, and its frequency will be double that of the fundamental.

With two extra nodes between the extremes the string will vibrate in three loops (Fig. 385; bottom). The wave length will be one-third that of the fundamental. Its frequency will be thrice that of the fundamental.

These notes, whose frequencies are integral multiples of the fundamental, are called *harmonics*.

353. Laws of Transverse Vibration of Strings

The frequency of the note emitted by a string, fixed at both ends and vibrating under tension, depends upon a number of factors. The nature of dependence is expressed by what are known as the laws of transverse vibration of strings. They may be stated as follows :

(i) Law of length. The frequency (n) varies inversely as the length (l) when the mass (m) per unit length and the tension (T) of the string remain constant.

In symbols, $n \propto 1/l$ when m and T are constant.

(ii) Law of tension. The frequency varies directly as the square root of the tension when the length and mass per unit length remain constant.

In symbols, $n \propto \sqrt{T}$ when l and m are constant.

(iii) Law of mass. The frequency varies inversely as the square root of the mass per unit length when the length and tension remain constant. Or, $n \propto \sqrt{1/m}$ when l and T are constant.

Since $m = \frac{1}{4}\pi d^2 \rho$ where d = diameter of the string and ρ=density of its material, we may split the law of mass into two laws as follows :

(iii) *(a)* *Law of diameter.* If the length, tension and material remain the same, the frequency varies inversely as the diameter.

(iii) *(b)* *Law of density.* If the length, tension and diameter are the same, the frequency varies inversely as the square root of the density of the material of the string.

It will be easy to remember these laws if we remember the formula

$$n = \frac{1}{2l}\sqrt{\frac{T}{m}}.$$

The formula is not difficult to establish, but we may leave it for a later stage. $\sqrt{T/m}$ is the velocity with which transverse waves move along a stretched string. l stands for the length of a vibrating loop, i.e., the distance between two consecutive nodes. n is the corresponding frequency.

354. Sonometer

The laws of transverse vibrations of strings can be verified with a sonometer. It consists of two wires stretched on a hollow, wooden box about a metre long (Fig. 386). One end of each of the wires is fixed to a pin. The other end passes over a smooth pulley and supports a suspended hook carrying weights, which supply the tension. It is however more convenient to dispense with the pulley in one of the wires and alter its tension by wrapping it round an iron peg which may be turned by means of a wrench key. Its tension can then change continuously. This wire may be called the auxiliary or reference wire and the other, the main wire. To vary the vibrating length at will, movable wooden bridges are placed below the wires. The wooden box vibrates in tune with the wires and greatly amplifies the sound.

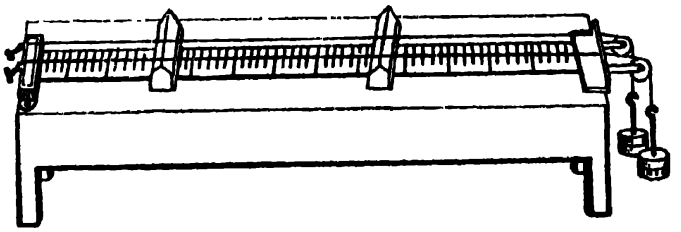


Fig. 386

If a stretched wire is plucked it vibrates with a definite frequency depending on the length, mass and tension of the vibrating segment. The frequency can be altered at will by varying the tension of the wire or by displacing any of the bridges, which alters the vibrating length.

Tuning of a sonometer. A person having a musical ear can very easily tune a length of the sonometer wire to the note from a vibrating tuning fork. But the following mechanical method may be used by all. A small piece of paper, called a *paper rider*, is bent in the form of an inverted V and placed in the middle of the wire. When the stem of the vibrating fork is pressed on the sonometer box, the vibrations are communicated to the wire through the wood. The string vibrates lightly at first. As the bridges are adjusted so that the frequency of the wire approaches that of the fork, the wire vibrates with increasing amplitude. Near resonance, the paper rider is thrown off the wire by the increased vibrations. Beats are heard if in this condition the fork and the wire are sounded together. The bridges are then so adjusted that beats disappear. This indicates perfect tuning.

Experiment 59. *To find the ratio of frequencies of two forks by a sonometer.*

Apparatus : Sonometer, a few half-kilogram weights, experimental forks, scale.

Method : Keep the sonometer wire under a fixed tension. The weight to be used will depend on the thickness of the wire. The wire should neither be slack nor under too high a tension.

Take one of the forks and bring a length of the sonometer wire into resonance with it, using the method of paper rider just described. Measure the resonant length.

Without altering the tension of the wire find, in the same way, the resonant length for the other fork.

If n_1 and n_2 are the frequencies of the forks, and l_1 and l_2 the respective resonant lengths, then

$$\frac{n_1}{n_2}=\frac{l_2}{l_1}.$$

The relation follows from the law of length of the transverse vibration of strings.

Change the tension to another value and repeat. The ratio of n_1 to n_2 will remain the same.

Verification of law of length. Similar observations will verify the law of length. A number of forks of known frequencies may be taken and the resonant length determined for each, keeping the tension fixed. It will be found that the product of the frequency and the corresponding resonant length is a constant. This shows that frequency and resonant length are inversely proportional.

355. Stringed Musical Instruments

The sitar, esraj, piano, violin, etc., are some of the well-known stringed musical instruments. They differ in the design as well as in the method by which the strings are set into vibration. The sitar is plucked with the fingers or picks, the piano is struck with light felt mallets, and the esraj and violin are bowed with long strands of tightly stretched horse-hair.

The strings on an instrument differ in length and thickness. The musician may, while playing, change the tension too. Thicker and heavier strings produce notes of lower pitch than thinner and lighter strings. The tighter a string is stretched the higher is its pitch.

The note emitted by the main string is strengthened by forced vibration of the board on which it is mounted. In some cases there is also resonance with other strings. (Also see § 345).

Vibration of Air Columns

Stationary waves (§ 348) are generated in a column of air when two identical wave trains move through it in opposite directions. Such a vibrating column of air can serve as a source of sound. Instruments like the organ, flute, clarinet etc. are designed on this property of air columns. The air column is enclosed in a pipe or tube. The shapes of these tubes and the methods by which the air in them is excited differ from instrument to instrument.

The pipe containing the air column may be closed at one end and open at the other, or open at both ends. In the former case we call it a *closed or stopped pipe* ; in the latter case it is called an *open pipe.*

In discussing vibrations of air columns we shall do well to remember the following facts :

(*i*) *The waves in the air column are stationary waves.*

(*ii*) *An open end of the pipe is always an antinode.*

(*iii*) *A closed end of the pipe is always a node.*

(*iv*) *The distance between a consecutive node and an antinode is always one fourth of the wave-length (of the stationary waves or of the sound waves emitted).*

(*v*) *Nodes and antinodes occur alternately along the pipe.*

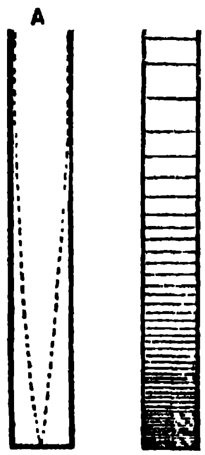
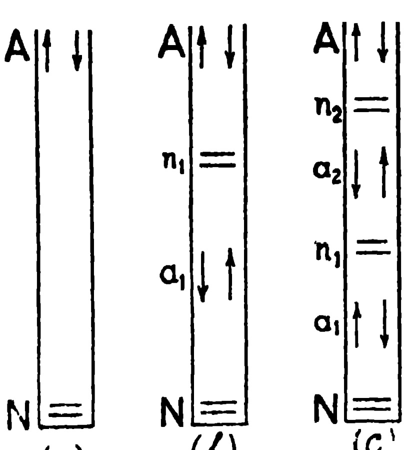


Fig. 387

A. Closed pipes. In the simplest mode of vibration the closed end is a node and the open end is its nearest antinode (Fig. 387 and 388 *a*); N, node : A, antinode). The length *L* of the pipe is, therefore, a quarter of the wave length of the note emitted. The note emitted is called the *fundamental.* Of all the notes the pipe can emit, the fundamental has the longest wave-length, and hence the lowest frequency.



(*a*) (*b*) (*c*)

Fig. 388

Fig. 388 (*b*) shows the next mode of vibration. In it there is only one pair of additional node and antinode (a_1, n_1) between the ends. The tube length now is equal to three-fourths of the wave-length emitted (since distance between a consecutive node and antinode is a quarter wave-length). The frequency of the emitted note is three times that of the fundamental.

Fig. 388 (*c*) shows the next higher mode of vibration. It has two pairs of additional nodes and antinodes (a_1, n_1, a_2, n_2) between the ends. The wave-length of the note emitted is one-fifth that of the fundamental and its frequency is five times the frequency of the fundamental. It is interesting to note that *in closed pipes it is possible to get only those harmonics whose frequencies are odd multiples of the fundamental.*

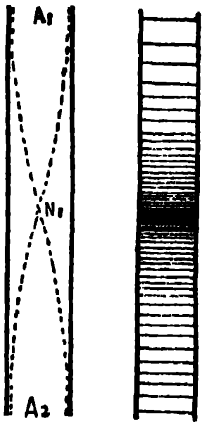


Fig. 389

B. Open pipes. In the simplest mode of vibration there is an antinode at each of the open ends with only one node between them (Fig. 389 and 390 *a*). This is the *fundamental.* Its wave-length is twice the length of the pipe. $\lambda = 2L$.

The modes of vibration giving rise to the next two harmonics are shown in figs. 390 (*b*) and (*c*). It is easy to see that these harmonics have frequencies respectively twice and thrice that of the fundamental. *The open pipe can produce harmonics with frequencies which may be odd as well as even multiples of the fundamental frequency.*

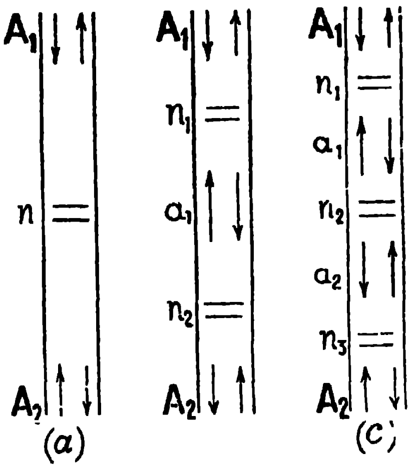


Fig. 390

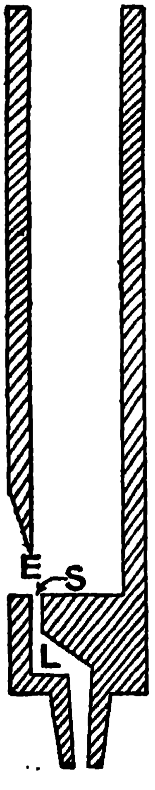


Fig. 391

End correction. Strictly speaking the antinode at an open end of a pipe is formed a little away from the end. Lord Rayleigh found that this distance is 0·3 times the diameter of the pipe. In any work involving measurement of wave-length of sound in pipes this correction shoud be considered. It is known as *end correction.*

357. Organ Pipes

Organ pipes are wooden or metal tubes and may be rectangular or circular in section. At one end of the pipe, a blast of air from the wind chest enters the mouth piece *L* (Fig. 391) and escapes through the slit *S*. It then impinges upon the bevelled edge *E* of a rectangular opening in the side of the pipe. This end of the pipe is to be treated as an open end. The other end of the pipe may be open (*open pipe*) or it may be closed by an adjustable piston (*stopped* or *closed pipe*).

The blast of air impinging on E sets the air column into vibration. The pitch of the note emitted by the pipe is determined by its length and by its 'open' or 'closed' nature. An open pipe has a fundamental frequency which is twice that of a closed pipe of the same length. The diameter or the width of the pipe has no effect on the pitch.

Closed organ pipes are tuned to the required frequency by shifting the adjustable piston it has at the closed end. Open pipes may have another sliding pipe outside to adjust its length. Alternatively, it may have a flap at one end partly covering the end. The frequency can be altered by changing the position of the flap.

358. The Phonograph

The phonograph, invented by Thomas Alva Edison in 1878, was the first device which successfully recorded and reproduced the sounds of speech and music. Its construction and action are as follows :

The sound to be recorded is produced in front of a horn (H : Fig. 392). The narrow end of the horn is closed by a circular diaphragm D. A light style P is fixed at right angles to D at its centre.

The sound waves, collected by the horn, cause the diaphragm to vibrate according to the variations of pressure in these waves. This causes the style to vibrate at right angles to the diaphragm.

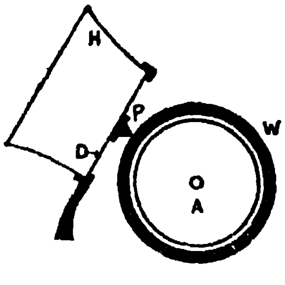


Fig. 392

The style is in contact with a layer W of special wax covering the drum A. This drum can rotate about its axis and advance as it rotates. If no sound waves impinge on H as the drum rotates and advances, the style traces a groove of uniform depth in the wax. But when sound waves fall on D, the style vibrates and traces a grove of varying depth. This is the *record* of the sound.

To reproduce the sound from this record, the style is placed at the starting point and the drum set in motion. It is made to move at the same constant speed as when taking the record. The point of the style follows the groove and moves up and down. This makes the diaphragm vibrate in the same way as it did at the time of recording. The air is set into similar vibration, reproducing the sound.

Since the invention of the phonograph there has been much improvement both in recording and reproduction of sound. The gramophone with disc records, the tape-recorder and sound-tracks on talkie-films are examples. The modern dictaphone has preserved some features of the phonograph.

Exercises

1. How can you establish that sound is due to the vibration of a body and that it requires a medium for propagation ?

2. Describe how sound is propagated through air.

3. Draw a neat diagram of the human ear showing clearly the structures which are essential for the transmission of sound to the inner ear. Label the parts, and state briefly how the ear acts.

4. Distinguish between transverse and longitudinal waves, giving an example of each.

Define the terms wave-length, frequency and velocity in reference to waves, and state the relation between them.

5. Describe a simple method of determining the speed of sound in open air. What are the common sources of error in such an experiment ? How are they minimized ?

6. What is an echo ? Can it be heard at all distances ? Say why. What causes reverberation in a hall and rumbling of thunder ?

7. What are supersonics ? Mention any use that you know of them.

8. Distinguish between free vibration and forced vibration. What is resonance ?

Describe the resonance tube method of determining the speed of sound. Is it the speed (*i*) in open air, (*ii*) in the air of the tube, or (*iii*) in the steel of the fork ?

9. What are beats ? Two forks produce five beats per second when sounded together. What will be the effects of loading a prong of (*a*) the fork of higher frequency, (*b*) the fork of lower frequency ?

10. What are stationary waves ? What are their characteristics ? Define the terms *nodes* and *antinodes*. What is the distance, in terms of wave-length, between (*i*) two consecutive nodes, (*ii*)

a consecutive node and antinode, (*iii*) two alternate nodes, (*iv*) two consecutive antinodes ?

11. Show with the help of neat diagrams how a tuning fork vibrates. What kind of motion does the stem execute ? Give an example where we utilize the motion of the stem.

12. What are the characteristics of a musical sound ? Mention in each case the physical feature of the vibrations with which a characteristic is associated.

What analogies do these characteristics have with light ?

13. What do you understand by the *quality* of a musical note ? Explain clearly the meanings of the terms *fundamental*, *harmonics* and *overtones* in this connection.

14. State the laws of transverse vibration of strings.

State also how a vibrating string produces the fundamental and also harmonics.

15. How would you use a sonometer (*i*) to demonstrate resonance, (*ii*) to compare the frequencies of two forks ?

16. You are given stringed musical instruments such as a sitar, an esraj and a violin, and asked to demonstrate (*i*) forced vibration, (*ii*) resonance and (*iii*) the laws of transverse vibration of strings.

Say what you will do.

17. Describe with neat diagrams how a closed pipe and an open pipe of the same length produce the fundamental. Also find the relation between the fundamental frequencies of the two.

How do the two kinds of pipe differ in the production of harmonics ?

18. Describe a simple device for the recording and reproduction of sound.